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ELEVENTH EDITION

VOLUME IX SLICE II

Ehud to Electroscope

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EHUD, in the Bible, a "judge" who delivered Israel from the Moabites (Judg. iii. 12-30). He was sent from Ephraim to bear tribute to Eglon king of Moab, who had crossed over the Jordan and seized the district around Jericho. Being, like the Benjamites, left-handed (cf. xx. 16), he was able to conceal a dagger and strike down the king before his intentions were suspected. He locked Eglon in his chamber and escaped. The men from Mt Ephraim collected under his leadership and by seizing the fords of the Jordan were able to cut off the Moabites. He is called the son of Gera a Benjamite, but since both Ehud and Gera are tribal names (2 Sam. xvi. 5, 1 Chron. viii. 3, 5 sq.) it has been thought that this notice is not genuine. The tribe of Benjamin rarely appears in the old history of the Hebrews before the time of Saul. See further BENJAMIN; JUDGES. 131

EIBENSTOCK, a town of Germany, in the kingdom of Saxony, near the Mulde, on the borders of Bohemia, 17 m. by rail S.S.E. of Zwickau. Pop. (1905) 7460. It is a principal seat of the tambour embroidery which was introduced in 1775 by Clara Angermann. It possesses chemical and tobacco manufactories, and tin and iron works. It has also a large cattle market. Eibenstock, together with Schwarzenberg, was acquired by purchase in 1533 by Saxony and was granted municipal rights in the following year.

EICHBERG, JULIUS (1824-1893), German musical composer, was born at Düsseldorf on the 13th of June 1824. When he was nineteen he entered the Brussels Conservatoire, where he took first prizes for violin-playing and composition. For eleven years he occupied the post of professor in the Conservatoire of Geneva. In 1857 he went to the United States, staying two years in New York and then proceeding to Boston, where he became director of the orchestra at the Boston Museum. In 1867 he founded the Boston Conservatory of Music. Eichberg published several educational works on music; and his four operettas, *The Doctor of Alcantara*, *The Rose of Tyrol*, *The Two Cadis* and *A Night in Rome*, were highly popular. He died in Boston on the 18th of January 1893.

EICHENDORFF, JOSEPH, FREIHERR VON (1788-1857), German poet and romance-writer, was born at Lubowitz, near Ratibor, in Silesia, on the 10th of March 1788. He studied law at Halle and Heidelberg from 1805 to 1808. After a visit to Paris he went to Vienna, where he resided until 1813, when he joined the Prussian army as a volunteer in the famous Lützow corps. When peace was concluded in 1815, he left the army, and in the following year he was appointed to a judicial office at Breslau. He subsequently held similar offices at Danzig, Königsberg and Berlin. Retiring from public service in 1844, he lived successively in Danzig, Vienna, Dresden and Berlin. He died at Neisse on the 26th of November 1857. Eichendorff was one of the most distinguished of the later members of the German romantic school. His genius was essentially lyrical. Thus he is most successful in his shorter romances and dramas, where constructive power is least called for. His first work, written in 1811, was a romance, *Ahnung und Gegenwart* (1815). This was followed at short intervals by several others, among which the foremost place is by general consent assigned to *Aus dem Leben eines Taugenichts* (1826), which has often been reprinted. Of his dramas may be mentioned *Ezzelin von Romano* (1828); and *Der letzte Held von Marienburg* (1830), both tragedies; and a comedy, *Die Freier* (1833). He also translated several of Calderon's religious dramas (*Geistliche Schauspiele*, 1846). It is, however, through his lyrics (*Gedichte*, first collected 1837) that Eichendorff is best known; he is the greatest lyric poet of the romantic movement. No one has given more beautiful expression than he to the poetry of a wandering life; often, again, his lyrics are exquisite word pictures interpreting the mystic meaning of the moods of nature, as in *Nachts*, or the old-time mystery which yet haunts the twilight forests and feudal castles of Germany, as in the dramatic lyric *Waldesgespräch* or *Auf einer Burg*. Their language is simple and musical, which makes them very suitable for singing, and they have been often set, notably by Schubert and Schumann.

In the later years of his life Eichendorff published several works on subjects in literary history and criticism such as *Über die ethische und religiöse Bedeutung der neuen romantischen Poesie in Deutschland* (1847), *Der deutsche Roman des 18. Jahrhunderts in seinem Verhältniss zum Christenthum* (1851), and *Geschichte der poetischen Litteratur Deutschlands* (1856), but the value of these works is impaired by the author's reactionary standpoint. An edition of his collected works in six volumes, appeared at Leipzig in 1870.

Eichendorff's *Sämtliche Werke* appeared in 6 vols., 1864 (reprinted 1869-1870); his *Sämtliche poetische Werke* in 4 vols. (1883). The latest edition is that edited by R. von Gottschall in 4 vols. (1901). A good selection edited by M. Kaoch will be found in vol. 145 of Kürschner's *Deutsche Nationalliteratur* (1893). Eichendorff's critical writings were collected in 1866 under the title *Vermischte Schriften* (5 vols.). Cp. H. von Eichendorff's biographical introduction to the *Sämtliche Werke*; also H. Keiter, *Joseph von Eichendorff* (Cologne, 1887); H.A. Krüger, *Der junge Eichendorff* (Oppeln, 1898).

EICHHORN, JOHANN GOTTFRIED (1752-1827), German theologian, was born at Dörrenzimmern, in the principality of Hohenlohe-Oehringen, on the 16th of October 1752. He was educated at the state school in Weikersheim, where his father was superintendent, at the gymnasium at Heilbronn and at the university of Göttingen (1770-1774), studying under J.D. Michaelis. In 1774 he received the rectorship of the gymnasium at Ohrdruf, in the duchy of Gotha, and in the following year was made professor of Oriental languages at Jena. On the death of Michaelis in 1788 he was elected professor *ordinarius* at Göttingen, where he lectured not only on Oriental languages and on the exegesis of the Old and New Testaments, but also on political history. His health was shattered in 1825, but he continued his lectures until attacked by fever on the 14th of June 1827. He died on the 27th of that month. Eichhorn has been called "the founder of modern Old Testament criticism." He first properly recognized its scope and problems, and began many of its most important discussions. "My greatest trouble," he says in the preface to the second edition of his *Einleitung*, "I had to bestow on a hitherto unworked field—on the investigation of the inner nature of the Old Testament with the help of the Higher Criticism (not a new name to any humanist)." His investigations led him to the conclusion that "most of the writings of the Hebrews have passed through several hands." He took for granted that all the so-called supernatural facts relating to the Old and New Testaments were explicable on natural principles. He sought to judge them from the standpoint of the ancient world, and to account for them by the superstitious beliefs which were then generally in vogue. He did not perceive in the biblical books any religious ideas of much importance for modern times; they interested him merely historically and for the light they cast upon antiquity. He regarded many books of the Old Testament as spurious, questioned the genuineness of *2 Peter* and *Jude*, denied the Pauline authorship of *Timothy* and *Titus*, and suggested that 132
the canonical gospels were based upon various translations and editions of a primary Aramaic gospel. He did not appreciate as sufficiently as David Strauss and the Tübingen critics the difficulties which a natural theory has to surmount, nor did he support his conclusions by such elaborate discussions as they deemed necessary.

His principal works were—*Geschichte des Ostindischen Handels vor Mohammed* (Gotha, 1775); *Allgemeine Bibliothek der biblischen Literatur* (10 vols., Leipzig, 1787-1801); *Einleitung in das Alte Testament* (3 vols., Leipzig, 1780-1783); *Einleitung in das Neue Testament* (1804-1812); *Einleitung in die apokryphischen Bücher des Alten Testaments* (Gött., 1795); *Commentarius in apocalypsin Joannis* (2 vols., Gött., 1791); *Die Hebr. Propheten* (3 vols., Gött., 1816-1819); *Allgemeine Geschichte der Cultur und Literatur des neuern Europa* (2 vols., Gött., 1796-1799); *Literärgeschichte* (1st vol., Gött., 1799, 2nd ed. 1813, 2nd vol. 1814); *Geschichte der Literatur von ihrem Anfange bis auf die neuesten Zeiten* (5 vols., Gött., 1805-1812); *Übersicht der Französischen Revolution* (2 vols., Gött., 1797); *Weltgeschichte* (3rd ed., 5 vols., Gött., 1819-1820); *Geschichte der drei letzten Jahrhunderte* (3rd ed., 6 vols., Hanover, 1817-1818); *Urgeschichte des erlauchten Hauses der Welfen* (Hanover, 1817).

See R.W. Mackay, *The Tübingen School and its Antecedents* (1863), pp. 103 ff.; Otto Pfleiderer, *Development of Theology* (1890), p. 209; T.K. Cheyne, *Founders of Old Testament Criticism* (1893), pp. 13 ff.

EICHHORN, KARL FRIEDRICH (1781-1854), German jurist, son of the preceding, was born at Jena on the 20th of November 1781. He entered the university of Göttingen in 1797. In 1805 he obtained the professorship of law at Frankfort-on-Oder, holding it till 1811, when he accepted the same chair at Berlin. On the call to arms in 1813 he became a captain of horse, and received at the end of the war the decoration of the Iron Cross. In 1817 he was offered the chair of law at Göttingen, and, preferring it to the Berlin professorship, taught there with great success till ill-health compelled him to resign in 1828. His successor in the Berlin chair having died in 1832, he again entered on its duties, but resigned two years afterwards. In 1832 he also received an appointment in the ministry of foreign affairs, which, with his labours on many state committees and his legal researches and writings, occupied him till his death at Cologne on the 4th of July 1854. Eichhorn is regarded as one of the principal authorities on German constitutional law. His chief work is *Deutsche Staats- und Rechtsgeschichte* (Göttingen, 1808-1823, 5th ed. 1843-1844). In company with Savigny and J.F.L. Göschen he founded the *Zeitschrift für geschichtliche Rechtswissenschaft*. He was the author besides of *Einleitung in das deutsche Privatrecht mit Einschluss des Lehnrechts* (Gött., 1823) and the *Grundsätze des Kirchenrechts der Katholischen und der Evangelischen Religionspartei in Deutschland*, 2 Bde. (*ib.*, 1831-1833).

See Schulte, *Karl Friedrich Eichhorn, sein Leben und Wirken* (1884).

EICHSTÄTT, a town and episcopal see of Germany, in the kingdom of Bavaria, in the deep and romantic valley of the Altmühl, 35 m. S. of Nuremberg, on the railway to Ingolstadt and Munich. Pop. (1905) 7701. The town, with its numerous spires and remains of medieval fortifications, is very picturesque. It has an Evangelical and seven Roman Catholic churches, among the latter the cathedral of St Wilibald (first bishop of Eichstätt),—with the tomb of the saint and numerous pictures and relics,—the church of St Walpurgis, sister of Wilibald, whose remains rest in the choir, and the Capuchin church, a copy of the Holy Sepulchre. Of its secular buildings the most noticeable are the town hall and the Leuchtenberg palace, once the residence of the prince bishops and later of the dukes of Leuchtenberg (now occupied by the court of justice of the district), with beautiful grounds. The Wilibaldsburg, built on a neighbouring hill in the 14th century by Bishop Bertold of Hohenzollern, was long the residence of the prince bishops of Eichstätt, and now contains an historical museum. There are an episcopal lyceum, a clerical seminary, a classical and a modern school, and numerous religious houses. The industries of the town include bootmaking, brewing and the production of lithographic stones.

Eichstätt (Lat. *Aureatum* or *Rubilocus*) was originally a Roman station which, after the foundation of the bishopric by Boniface in 745, developed into a considerable town, which was surrounded with walls in 908. The bishops of Eichstätt were princes of the Empire, subject to the spiritual jurisdiction of the archbishops of Mainz, and ruled over considerable territories in the Circle of Franconia. In 1802 the see was secularized and incorporated in Bavaria. In 1817 it was given, with the duchy of Leuchtenberg, as a mediatized domain under the Bavarian crown, by the king of Bavaria to his son-in-law Eugène de Beauharnais, ex-viceroy of Italy, henceforth styled duke of Leuchtenberg. In 1855 it reverted to the Bavarian crown.

EICHWALD, KARL EDUARD VON (1795-1876), Russian geologist and physician, was born at Mitau in Courland on the 4th of July 1795. He became doctor of medicine and professor of zoology in Kazañ in 1823; four years later professor of zoology and comparative anatomy at Vilna; in 1838 professor of zoology, mineralogy and medicine at St Petersburg; and finally professor of palaeontology in the institute of mines in that city. He travelled much in the Russian empire, and was a keen observer of its natural history and geology. He died at St Petersburg on the 10th of November 1876. His published works include *Reise auf dem Caspischen Meere und in den Caucasus*, 2 vols. (Stuttgart and Tübingen, 1834-1838); *Die Urwelt Russlands* (St Petersburg, 1840-1845); *Lethaea Rossica, ou paléontologie de la Russie*, 3 vols. (Stuttgart, 1852-1868), with Atlases.

EIDER, a river of Prussia, in the province of Schleswig-Holstein. It rises to the south of Kiel, in Lake Redder, flows first north, then west (with wide-sweeping curves), and after a course of 117 m. enters the North Sea at Tönning. It is navigable up to Rendsburg, and is embanked through the marshes across which it runs in its lower course. Since the reign of Charlemagne, the Eider (originally *Ägyr Dör*—Neptune’s gate) was known as *Romani terminus imperii* and was recognized as the boundary of the Empire in 1027 by the emperor Conrad II., the founder of the Salian dynasty. In the controversy arising out of the Schleswig-Holstein Question, which culminated in the war of Austria and Prussia against Denmark in 1864, the Eider gave its name to the “Eider Danes,” the *intransigeant* Danish party which maintained that Schleswig (Sonderjylland, South Jutland) was by nature and historical tradition an integral part of Denmark. The Eider Canal (*Eider-Kanal*), which was constructed between 1777 and 1784, leaves the Eider at the point where the river turns to the west and enters the Bay of Kiel at Holtenau. It was hampered by six sluices, but was used annually by some 4000 vessels, and until its conversion in 1887-1895 into the Kaiser Wilhelm Canal afforded the only direct connexion between the North Sea and the Baltic.

EIDER (Icelandic, *Ædur*), a large marine duck, the *Somateria mollissima* of ornithologists, famous for its down, which, from its extreme lightness and elasticity, is in great request for filling bed-coverlets. This bird generally frequents low rocky islets near the coast, and in Iceland and Norway has long been afforded every encouragement and protection, a fine being inflicted for killing it during the breeding-season, or even for firing a gun near its haunts, while artificial nesting-places are in many localities contrived for its further accommodation. From the care thus taken of it in those countries it has become exceedingly tame at its chief resorts, which are strictly regarded as property, and the taking of eggs or down from them, except by authorized persons, is severely punished by law. In appearance the eider is somewhat clumsy, though it flies fast and dives admirably. The female is of a dark reddish-brown colour barred with brownish-black. The adult male in spring is conspicuous by his pied plumage of velvet-black beneath, and white above: a patch of shining sea-green on his head is only seen on close inspection. This plumage he is considered not to acquire until his third year, being when young almost exactly like the female, and it is certain that the birds which have not attained their full dress remain in flocks by themselves without going to the breeding-stations. The nest is generally in some convenient corner among large stones, hollowed in the soil, and furnished with a few bits of dry grass, seaweed or heather. By the time that the full number of eggs (which rarely if ever exceeds five) is laid the down is added. Generally the eggs and down are taken at intervals of a few days by 133
the owners of the "eider-fold," and the birds are thus kept depositing both during the whole season; but some experience is needed to ensure the greatest profit from each commodity. Every duck is ultimately allowed to hatch an egg or two to keep up the stock, and the down of the last nest is gathered after the birds have left the spot. The story of the drake's furnishing down, after the duck's supply is exhausted is a fiction. He never goes near the nest. The eggs have a strong flavour, but are much relished by both Icelanders and Norwegians. In the Old World the eider breeds in suitable localities from Spitsbergen to the Farne Islands off the coast of Northumberland—where it is known as St Cuthbert's duck. Its food consists of marine animals (molluscs and crustaceans), and hence the young are not easily reared in captivity. The eider of the New World differs somewhat, and has been described as a distinct species (*S. dresseri*). Though much diminished in numbers by persecution, it is still abundant on the coast of Newfoundland and thence northward. In Greenland also eiders are very plentiful, and it is supposed that three-fourths of the supply of down sent to Copenhagen comes from that country. The limits of the eider's northern range are not known, but the Arctic expedition of 1875 did not meet with it after leaving the Danish settlements, and its place was taken by an allied species, the king-duck (*S. spectabilis*), a very beautiful bird which sometimes appears on the British coast. The female greatly resembles that of the eider, but the male has a black chevron on his chin and a bright orange prominence on his forehead, which last seems to have given the species its English name. On the west coast of North America the eider is represented by a species (*S. v-nigrum*) with a like chevron, but otherwise resembling the Atlantic bird. In the same waters two other fine species are also found (*S. fischeri* and *S. stelleri*), one of which (the latter) also inhabits the Arctic coast of Russia and East Finmark and has twice reached England. The Labrador duck (*S. labradoria*), now extinct, also belongs to this group.

(A. N.)

EIFEL, a district of Germany, in the Prussian Rhine Province, between the Rhine, the Moselle and the frontier of the grand duchy of Luxemburg. It is a hilly region, most elevated in the eastern part (Hohe Eifel), where there are several points from 2000 up to 2410 ft. above sea-level. In the west is the Schneifels or Schnee-Eifel; and the southern part, where the most picturesque scenery and chief geological interest is found, is called the Vorder Eifel.

The Eifel is an ancient massif of folded Devonian rocks upon the margins of which, near Hillesheim and towards Bitburg and Trier, rest unconformably the nearly undisturbed sandstones, marls and limestones of the Trias. On the southern border, at Wittlich, the terrestrial deposits of the Permian Rothliegende are also met with. The slates and sandstones of the Lower Devonian form by far the greater part of the region; but folded amongst these, in a series of troughs running from south-west to north-east lie the fossiliferous limestones of the Middle Devonian, and occasionally, as for example near Büdesheim, a few small patches of the Upper Devonian. Upon the ancient floor of folded Devonian strata stand numerous small volcanic cones, many of which, though long extinct, are still very perfect in form. The precise age of the eruptions is uncertain. The only sign of any remaining volcanic activity is the emission in many places of carbon dioxide and of heated waters. There is no historic or legendary record of any eruption, but nevertheless the eruptions must have continued to a very recent geological period. The lavas of Papenkaule are clearly posterior to the excavation of the valley of the Kyll, and an outflow of basalt has forced the Uess to seek a new course. The volcanic rocks occur both as tuffs and as lava-flows. They are chiefly leucite and nepheline rocks, such as leucitite, leucitophyre and nephelinite, but basalt and trachyte also occur. The leucite lavas of Niedermendig contain haüyne in abundance. The most extensive and continuous area of volcanic rocks is that surrounding the Laacher See and extending eastwards to Neuwied and Coblenz and even beyond the Rhine.

The numerous so-called crater-lakes or *maare* of the Eifel present several features of interest. They do not, as a rule, lie in true craters at the summit of volcanic cones, but rather in hollows which have been formed by explosions. The most remarkable group is that of Daun, where the three depressions of Gemünd, Weinfeld and Schalkenmehren have been hollowed out in the Lower Devonian strata. The first of these shows no sign of either lavas or scoriae, but volcanic rocks occur on the margins of the other two. The two largest lakes in the Eifel region, however, are the Laacher See in the hills west of Andernach on the Rhine, and the Pulvermaar S.E. of the Daun group, with its shores of peculiar volcanic sand, which also appears in its waters as a black powder (*pulver*).

EIFFEL TOWER. Erected for the exposition of 1889, the Eiffel Tower, in the Champ de Mars, Paris, is by far the highest artificial structure in the world, and its height of 300 metres (984 ft.) surpasses that of the obelisk at Washington by 429 ft., and that of St Paul’s cathedral by 580 ft. Its framework is composed essentially of four uprights, which rise from the corners of a square measuring 100 metres on the side; thus the area it covers at its base is nearly 2½ acres. These uprights are supported on huge piers of masonry and concrete, the foundations for which were carried down, by the aid of iron caissons and compressed air, to a depth of about 15 metres on the side next the Seine, and about 9 metres on the other side. At first they curve upwards at an angle of 54°; then they gradually become straighter, until they unite in a single shaft rather more than half-way up. The first platform, at a height of 57 metres, has an area of 5860 sq. yds., and is reached either by staircases or lifts. The next, accessible by lifts only, is 115 metres up, and has an area of 32 sq. yds; while the third, at 276, supports a pavilion capable of holding 800 persons. Nearly 25 metres higher up still is the lantern, with a gallery 5 metres in diameter. The work of building this structure, which is mainly composed of iron lattice-work, was begun on the 28th of January 1887, and the full height was reached on the 13th of March 1889. Besides being one of the sights of Paris, to which visitors resort in order to enjoy the extensive view that can be had from its higher galleries on a clear day, the tower is used to some extent for scientific and semi-scientific purposes; thus meteorological observations are carried on. The engineer under whose direction the tower was constructed was Alexandre Gustave Eiffel (born at Dijon on the 15th of December 1832), who had already had a wide experience in the construction of large metal bridges, and who designed the huge sluices for the Panama Canal, when it was under the French company.

EILDON HILLS, a group of three conical hills, of volcanic origin, in Roxburghshire, Scotland, 1 m. S. by E. of Melrose, about equidistant from Melrose and St Boswells stations on the North British railway. They were once known as Eldune—the *Eldunum* of Simeon of Durham (fl. 1130)—probably derived from the Gaelic *aill*, “rock,” and *dun*, “hill”; but the name is also said to be a corruption of the Cymric *moeldun*, “bald hill.” The northern peak is 1327 ft. high, the central 1385 ft. and the southern 1216 ft. Whether or not the Roman station of *Trimontium* was situated here is matter of controversy. According to General William Roy (1726-1790) Trimontium—so called, according to this theory, from the triple Eildon heights—was Old Melrose; other authorities incline to place the station on the northern shore of the Solway Firth. The Eildons have been the subject of much legendary lore. Michael Scot (1175-1234), acting as a confederate of the Evil One (so the fable runs) cleft Eildon Hill, then a single cone, into the three existing peaks. Another legend states that Arthur and his knights sleep in a vault beneath the Eildons. A third legend centres in Thomas of Erceldoune. The Eildon Tree Stone, a large moss-covered boulder, lying on the high road as it bends towards the west within 2 m. of Melrose, marks the spot where the Fairy Queen led him into her realms in the heart of the hills. Other places associated with this legend may still be identified. Huntly Banks, where “true Thomas” lay and watched the queen’s approach, is half a mile west of the Eildon Tree Stone, and on the west 134
side of the hills is Bogle Burn, a streamlet that feeds the Tweed and probably derives its name from his ghostly visitor. Here, too, is Rhymer’s glen, although the name was invented by Sir Walter Scott, who added the dell to his Abbotsford estate. Bowden, to the south of the hills, was the birthplace of the poets Thomas Aird (1802-1876) and James Thomson, and its parish church contains the burial-place of the dukes of Roxburghe. Eildon Hall is a seat of the duke of Buccleuch.

EILENBURG, a town of Germany, in the Prussian province of Saxony, on an island formed by the Mulde, 31 m. E. from Halle, at the junction of the railways Halle-Cottbus and Leipzig-Eilenburg. Pop. (1905) 15,145. There are three churches, two Evangelical and one Roman Catholic. The industries of the town include the manufacture of chemicals, cloth, quilting, calico, cigars and agricultural implements, bleaching, dyeing, basket-making, carriage-building and trade in cattle. In the neighbourhood is the iron foundry of Erwinhof. Opposite the town, on the steep left bank of the Mulde, is the castle from which it derives its name, the original seat of the noble family of Eulenburg. This castle (Ilburg) is mentioned in records of the reigns of Henry the Fowler as an important outpost against the Sorbs and Wends. The town itself, originally called Mildenau, is of great antiquity. It is first mentioned as a town in 981, when it belonged to the house of Wettin and was the chief town of the East Mark. In 1386 it was incorporated in the margraviate of Meissen. In 1815 it passed to Prussia.

See Gundermann, *Chronik der Stadt Eilenburg* (Eilenburg, 1879).

EINBECK, or Eimbeck, a town of Germany, in the Prussian province of Hanover, on the Ilm, 50 m. by rail S. of Hanover. Pop. (1905) 8709. It is an old-fashioned town with many quaint wooden houses, notable among them the "Northeimhaus," a beautiful specimen of medieval architecture. There are several churches, among them the Alexanderkirche, containing the tombs of the princes of Grubenhagen, and a synagogue. The schools include a *Realgymnasium* (*i.e.* predominantly for "modern" subjects), technical schools for the advanced study of machine-making, for weaving and for the textile industries, a preparatory training-college and a police school. The industries include brewing, weaving and the manufacture of cloth, carpets, tobacco, sugar, leather-grease, toys and roofing-felt.

Einbeck grew up originally round the monastery of St Alexander (founded 1080), famous for its relic of the True Blood. It is first recorded as a town in 1274, and in the 14th century was the seat of the princes of Grubenhagen, a branch of the ducal house of Brunswick. The town subsequently joined the Hanseatic League. In the 15th century it became famous for its beer ("Eimbecker," whence the familiar "Bock"). In 1540 the Reformation was introduced by Duke Philip of Brunswick-Saltzderhelden (d. 1551), with the death of whose son Philip II. (1596) the Grubenhagen line became extinct. In 1626, during the Thirty Years' War, Einbeck was taken by Pappenheim and in October 1641 by Piccolomini. In 1643 it was evacuated by the Imperialists. In 1761 its walls were razed by the French.

See H.L. Harland, *Gesch. der Stadt Einbeck*, 2 Bde. (Einbeck, 1854-1859; abridgment, *ib.* 1881).

EINDHOVEN, a town in the province of North Brabant, Holland, and a railway junction 8 m. by rail W. by S. of Helmond. Pop. (1900) 4730. Like Tilburg and Helmond it has developed in modern times into a flourishing industrial centre, having linen, woollen, cotton, tobacco and cigar, matches, &c., factories and several breweries.

EINHARD (*c.* 770-840), the friend and biographer of Charlemagne; he is also called Einhartus, Ainhardus or Heinhardus, in some of the early manuscripts. About the 10th century the name was altered into Agenardus, and then to Eginhardus, or Eginhartus, but, although these variations were largely used in the English and French languages, the form Einhardus, or Einhartus, is unquestionably the right one.

According to the statement of Walafrid Strabo, Einhard was born in the district which is watered by the river Main, and his birth has been fixed at about 770. His parents were of noble birth, and were probably named Einhart and Engilfrit; and their son was educated in the monastery of Fulda, where he was certainly residing in 788 and in 791. Owing to his intelligence and ability he was transferred, not later than 796, from Fulda to the palace of Charlemagne by abbot Baugulf; and he soon became very intimate with the king and his family, and undertook various important duties, one writer calling him *domesticus palatii regalis*. He was a member of the group of scholars who gathered around Charlemagne and was entrusted with the charge of the public buildings, receiving, according to a fashion then prevalent, the scriptural name of Bezaleel (Exodus xxxi. 2 and xxxv. 30-35) owing to his artistic skill. It has been supposed that he was responsible for the erection of the basilica at Aix-la-Chapelle, where he resided with the emperor, and the other buildings mentioned in chapter xvii. of his *Vita Karoli Magni*, but there is no express statement to this effect. In 806 Charlemagne sent him to Rome to obtain the signature of Pope Leo III. to a will which he had made concerning the division of his empire; and it was possibly owing to Einhard's influence that in 813, after the death of his two elder sons, the emperor made his remaining son, Louis, a partner with himself in the imperial dignity. When Louis became sole emperor in 814 he retained his father's minister in his former position; then in 817 made him tutor to his son, Lothair, afterwards the emperor Lothair I.; and showed him many other marks of favour. Einhard married Emma, or Imma, a sister of Bernharius, bishop of Worms, and a tradition of the 12th century represented this lady as a daughter of Charlemagne, and invented a romantic story with regard to the courtship which deserves to be noticed as it frequently appears in literature. Einhard is said to have visited the emperor's daughter regularly and secretly, and on one occasion a fall of snow made it impossible for him to walk away without leaving footprints, which would lead to his detection. This risk, however, was obviated by the foresight of Emma, who carried her lover across the courtyard of the palace; a scene which was witnessed by Charlemagne, who next morning narrated the occurrence to his counsellors, and asked for their advice. Very severe punishments were suggested for the clandestine lover, but the emperor rewarded the devotion of the pair by consenting to their marriage. This story is, of course, improbable, and is further discredited by the fact that Einhard does not mention Emma among the number of Charlemagne's children. Moreover, a similar story has been told of a daughter of the emperor Henry III. It is uncertain whether Einhard had any children. He addressed a letter to a person named Vussin, whom he calls *fili* and *mi nate*, but, as Vussin is not mentioned in documents in which his interests as Einhard's son would have been concerned, it is possible that he was only a young man in whom he took a special interest. In January 815 the emperor Louis I. bestowed on Einhard and his wife the domains of Michelstadt and Mulinheim in the Odenwald, and in the charter conveying these lands he is called simply Einhardus, but, in a document dated the 2nd of June of the same year, he is referred to as abbot. After this time he is mentioned as head of several monasteries: St Peter, Mount Blandin and St Bavon at Ghent, St Servais at Maastricht, St Cloud near Paris, and Fontenelle near Rouen, and he also had charge of the church of St John the Baptist at Pavia.

During the quarrels which took place between Louis I. and his sons, in consequence of the emperor's

second marriage, Einhard’s efforts were directed to making peace, but after a time he grew tired of the troubles and intrigues of court life. In 818 he had given his estate at Michelstadt to the abbey of Lorsch, but he retained Mulinheim, where about 827 he founded an abbey and erected a church, to which he transported some relics of St Peter and St Marcellinus, which he had procured from Rome. To Mulinheim, which was afterwards called Seligenstadt, he finally retired in 830. His wife, who had been his constant helper, and whom he had not put away on becoming an abbot, died in 836, and after receiving a visit from the emperor, Einhard died on the 14th of March 840. He was buried at Seligenstadt, and his epitaph was written by Hrabanus Maurus. Einhard was a man of very short stature, a feature on which Alcuin wrote an epigram. 135
Consequently he was called *Nardulus*, a diminutive form of Einhardus, and his great industry and activity caused him to be likened to an ant. He was also a man of learning and culture. Reaping the benefits of the revival of learning brought about by Charlemagne, he was on intimate terms with Alcuin, was well versed in Latin literature, and knew some Greek. His most famous work is his *Vita Karoli Magni*, to which a prologue was added by Walafrid Strabo. Written in imitation of the *De vitis Caesarum* of Suetonius, this is the best contemporary account of the life of Charlemagne, and could only have been written by one who was very intimate with the emperor and his court. It is, moreover, a work of some artistic merit, although not free from inaccuracies. It was written before 821, and having been very popular during the middle ages, was first printed at Cologne in 1521. G.H. Pertz collated more than sixty manuscripts for his edition of 1829, and others have since come to light. Other works by Einhard are: *Epistolae*, which are of considerable importance for the history of the times; *Historia translationis beatorum Christi martyrum Marcellini et Petri*, which gives a curious account of how the bones of these martyrs were stolen and conveyed to Seligenstadt, and what miracles they wrought; and *De adoranda cruce*, a treatise which has only recently come to light, and which has been published by E. Dümmler in the *Neues Archiv der Gesellschaft für ältere deutsche Geschichtskunde*, Band xi. (Hanover, 1886). It has been asserted that Einhard was the author of some of the Frankish annals, and especially of part of the annals of Lorsch (*Annales Laurissenses majores*), and part of the annals of Fulda (*Annales Fuldenses*). Much discussion has taken place on this question, and several of the most eminent of German historians, Ranke among them, have taken part therein, but no certain decision has been reached.

The literature on Einhard is very extensive, as nearly all those who deal with Charlemagne, early German and early French literature, treat of him. Editions of his works are by A. Teulet, *Einhardi omnia quae extant opera* (Paris, 1840-1843), with a French translation; P. Jaffé, in the *Bibliotheca rerum Germanicarum*, Band iv. (Berlin, 1867); G.H. Pertz in the *Monumenta Germaniae historica*, Bände i. and ii. (Hanover, 1826-1829), and J.P. Migne in the *Patrologia Latina*, tomes 97 and 104 (Paris, 1866). The *Vita Karoli Magni*, edited by G.H. Pertz and G. Waitz, has been published separately (Hanover, 1880). Among the various translations of the *Vita* may be mentioned an English one by W. Glaister (London, 1877) and a German one by O. Abel (Leipzig, 1893). For a complete bibliography of Einhard, see A. Potthast, *Bibliotheca historica*, pp. 394-397 (Berlin, 1896), and W. Wattenbach, *Deutschlands Geschichtsquellen*, Band i. (Berlin, 1904).

(A. W. H.*)

EINHORN, DAVID (1809-1879), leader of the Jewish reform movement in the United States of America, was born in Bavaria. He was a supporter of the principles of Abraham Geiger (*q.v.*), and while still in Germany advocated the introduction of prayers in the vernacular, the exclusion of nationalistic hopes from the synagogue service, and other ritual modifications. In 1855 he migrated to America, where he became the acknowledged leader of reform, and laid the foundation of the régime under which the mass of American Jews (excepting the newly arrived Russians) now worship. In 1858 he published his revised prayer book, which has formed the model for all subsequent revisions. In 1861 he strongly supported the anti-slavery party, and was forced to leave Baltimore where he then ministered. He continued his work first in Philadelphia and later in New York.

(I. A.)

EINSIEDELN, the most populous town in the Swiss canton of Schwyz. It is built on the right bank of the Alpbach (an affluent of the Sihl), at a height of 2908 ft. above the sea-level on a rather bare moorland, and by rail is 25 m. S.E. of Zürich, or by a round-about railway route about 38 m. north of Schwyz, with which it communicates directly over the Hacken Pass (4649 ft.) or the Holzegg Pass (4616 ft.). In 1900 the population was 8496, all (save 75) Romanists and all (save 111) German-speaking. The town is entirely dependent on the great Benedictine abbey that rises slightly above it to the east. Close to its present site Meinrad, a hermit, was murdered in 861 by two robbers, whose crime was made known by Meinrad’s two pet ravens. Early in the 10th century Benno, a hermit, rebuilt the holy man’s cell, but the abbey proper was not founded till about 934, the church having been consecrated (it is said by Christ Himself) in 948. In 1274 the dignity of a prince of the Holy Roman Empire was confirmed by the emperor to the reigning abbot. Originally under the protection of the counts of Rapperswil (to which town on the lake of Zürich the old pilgrims’ way still leads over the Etzel Pass, 3146 ft., with its chapel and inn), this position passed by marriage with their heiress in 1295 to the Laufenburg or cadet line of the Habsburgs, but from 1386 was permanently occupied by Schwyz. A black wooden image of the Virgin and the fame of St Meinrad caused the throngs of pilgrims to resort to Einsiedeln in the middle ages, and even now it is much frequented, particularly about the 14th of September. The existing buildings date from the 18th century only, while the treasury and the library still contain many precious objects, despite the sack by the French in 1798. There are now about 100 fully professed monks, who direct several educational institutions. The Black Virgin has a special chapel in the stately church. Zwingli was the parish priest of Einsiedeln 1516-1518 (before he became a Protestant), while near the town Paracelsus (1493-1541), the celebrated philosopher, was born.

See Father O. Ringholz, *Geschichte d. fürstl. Benediktinerstiftes Einsiedeln*, vol. i. (to 1526), (Einsiedeln, 1904).

(W. A. B. C.)

EISENACH, a town of Germany, second capital of the grand-duchy of Saxe-Weimar-Eisenach, lies at the north-west foot of the Thuringian forest, at the confluence of the Nesse and Hörsel, 32 m. by rail W. from Erfurt. Pop. (1905) 35,123. The town mainly consists of a long street, running from east to west. Off this are the market square, containing the grand-ducal palace, built in 1742, where the duchess Hélène of Orleans long resided, the town-hall, and the late Gothic St Georgenkirche; and the square on which stands the Nikolaikirche, a fine Romanesque building, built about 1150 and restored in 1887. Noteworthy are also the Klemda, a small castle dating from 1260; the Lutherhaus, in which the reformer stayed with the Cotta family in 1498; the house in which Sebastian Bach was born, and that (now a museum) in which Fritz Reuter lived (1863-1874). There are monuments to the two former in the town, while the resting-place of the latter in the cemetery is marked by a less pretentious memorial. Eisenach has a school of forestry, a school of design, a classical school (*Gymnasium*) and modern school (*Realgymnasium*), a deaf and dumb school, a teachers' seminary, a theatre and a Wagner museum. The most important industries of the town are worsted-spinning, carriage and wagon building, and the making of colours and pottery. Among others are the manufacture of cigars, cement pipes, iron-ware and machines, alabaster ware, shoes, leather, &c., cabinet-making, brewing, granite quarrying and working, tile-making, and saw- and corn-milling.

The natural beauty of its surroundings and the extensive forests of the district have of late years attracted many summer residents. Magnificently situated on a precipitous hill, 600 ft. above the town to the south, is the historic Wartburg (*q.v.*), the ancient castle of the landgraves of Thuringia, famous as the scene of the contest of Minnesingers immortalized in Wagner's Tannhäuser, and as the place where Luther, on his return from the diet of Worms in 1521, was kept in hiding and made his translation of the Bible. On a high rock adjacent to the Wartburg are the ruins of the castle of Mädelstein.

Eisenach (*Isenacum*) was founded in 1070 by Louis II. the Springer, landgrave of Thuringia, and its history during the middle ages was closely bound up with that of the Wartburg, the seat of the landgraves. The Klemda, mentioned above, was built by Sophia (d. 1284), daughter of the landgrave Louis IV., and wife of Duke Henry II. of Brabant, to defend the town against Henry III., margrave of Meissen, during the succession contest that followed the extinction of the male line of the Thuringian landgraves in 1247. The principality of Eisenach fell to the Saxon house of Wettin in 1440, and in the partition of 1485 formed part of the territories given to the Ernestine line. It was a separate Saxon duchy from 1596 to 1638, from 1640 to 1644, and again 136
from 1662 to 1741, when it finally fell to Saxe-Weimar. The town of Eisenach, by reason of its associations, has been a favourite centre for the religious propaganda of Evangelical Germany, and since 1852 it has been the scene of the annual conference of the German Evangelical Church, known as the Eisenach conference.

See Trinius, *Eisenach und Umgebung* (Minden, 1900); and H.A. Daniel, *Deutschland* (Leipzig, 1895), and further references in U. Chevalier, "Répertoire des sources," &c., *Topo-bibliogr.* (Montbéliard, 1894-1899), s.v.

EISENBERG (*Isenberg*), a town of Germany, in the duchy of Saxe-Altenburg, on a plateau between the rivers Saale and Elster, 20 m. S.W. from Zeitz, and connected with the railway Leipzig-Gera by a branch to Crossen. Pop. (1905) 8824. It possesses an old castle, several churches and monuments to Duke Christian of Saxe-Eisenberg (d. 1707), Bismarck, and the philosopher Karl Christian Friedrich Krause (*q.v.*). Its principal industries are weaving, and the manufacture of machines, ovens, furniture, pianos, porcelain and sausages.

See Back, *Chronik der Sladt und des Amtes Eisenberg* (Eisenb., 1843).

EISENERZ ("Iron ore"), a market-place and old mining town in Styria, Austria, 68 m. N.W. of Graz by rail. Pop. (1900) 6494. It is situated in a deep valley, dominated on the east by the Pfaffenstein (6140 ft.), on the west by the Kaiserschild (6830 ft.), and on the south by the Erzberg (5030 ft.). It has an interesting example of a medieval fortified church, a Gothic edifice founded by Rudolph of Habsburg in the 13th century and rebuilt in the 16th. The Erzberg or Ore Mountain furnishes such rich ore that it is quarried in the open air like stone, in the summer months. There is documentary evidence of the mines having been worked as far back as the 12th century. They afford employment to two or three thousand hands in summer and about half as many in winter, and yield some 800,000 tons of iron per annum. Eisenerz is connected with the mines by the Erzberg railway, a bold piece of engineering work, 14 m. long, constructed on the Abt's rack-and-pinion system. It passes through some beautiful scenery, and descends to Vordernberg (pop. 3111), an important centre of the iron trade situated on the south side of the Erzberg. Eisenerz possesses, in addition, twenty-five furnaces, which produce iron, and particularly steel, of exceptional excellence. A few miles to the N.W. of Eisenerz lies the castle of Leopoldstein, and near it the beautiful Leopoldsteiner Lake. This lake, with its dark-green water, situated at an altitude of 2028 ft., and surrounded on all sides by high peaks, is not big, but is very deep, having a depth of 520 ft.

EISLEBEN (Lat. *Islebia*), a town of Germany, in the Prussian province of Saxony, 24 m. W. by N. from Halle, on the railway to Nordhausen and Cassel. Pop. (1905) 23,898. It is divided into an old and a new town (Altstadt and Neustadt). Among its principal buildings are the church of St Andrew (Andreaskirche), which contains numerous monuments of the counts of Mansfeld; the church of St Peter and St Paul (Peter-Paulkirche), containing the font in which Luther was baptized; the royal gymnasium (classical school), founded by Luther shortly before his death in 1546; and the hospital. Eisleben is celebrated as the place where Luther was born and died. The house in which he was born was burned in 1689, but was rebuilt in 1693 as a free school for orphans. This school fell into decay under the régime of the kingdom of Westphalia, but was restored in 1817 by King Frederick William III. of Prussia, who, in 1819, transferred it to a new building behind the old house. The house in which Luther died was restored towards the end of the 19th century, and his death chamber is still preserved. A bronze statue of Luther by Rudolf Siemering (1835-1905) was unveiled in 1883. Eisleben has long been the centre of an important mining district (Luther was a miner’s son), the principal products being silver and copper. It possesses smelting works and a school of mining.

The earliest record of Eisleben is dated 974. In 1045, at which time it belonged to the counts of Mansfeld, it received the right to hold markets, coin money, and levy tolls. From 1531 to 1710 it was the seat of the cadet line of the counts of Mansfeld-Eisleben. After the extinction of the main line of the counts of Mansfeld, Eisleben fell to Saxony, and, in the partition of Saxony by the congress of Vienna in 1815, was assigned to Prussia.

See G. Grössler, *Urkundliche Gesch. Eislebens bis zum Ende des 12. Jahrhunderts* (Halle, 1875); *Chronicon Islebiense; Eisleben Stadtchronik aus den Jahren* 1520-1738, edited from the original, with notes by Grössler and Sommer (Eisleben, 1882).

EISTEDDFOD (plural Eisteddfodau), the national bardic congress of Wales, the objects of which are to encourage bardism and music and the general literature of the Welsh, to maintain the Welsh language and customs of the country, and to foster and cultivate a patriotic spirit amongst the people. This institution, so peculiar to Wales, is of very ancient origin.[1] The term *Eisteddfod*, however, which means “a session” or “sitting,” was probably not applied to bardic congresses before the 12th century.

The Eisteddfod in its present character appears to have originated in the time of Owain ap Maxen Wledig, who at the close of the 4th century was elected to the chief sovereignty of the Britons on the departure of the Romans. It was at this time, or soon afterwards, that the laws and usages of the Gorsedd were codified and remodelled, and its motto of “Y gwir yn erbyn y byd” (The truth against the world) given to it. “Chairs” (with which the Eisteddfod as a national institution is now inseparably connected) were also established, or rather perhaps resuscitated, about the same time. The chair was a kind of convention where disciples were trained, and bardic matters discussed preparatory to the great Gorsedd, each chair having a distinctive motto. There are now existing four chairs in Wales,—namely, the “royal” chair of Powys, whose motto is “A laddo a leddir” (He that slayeth shall be slain); that of Gwent and Glamorgan, whose motto is “Duw a phob daioni” (God and all goodness); that of Dyfed, whose motto is “Calon wrth galon” (Heart with heart); and that of Gwynedd, or North Wales, whose motto is “Iesu,” or “O Iesu! na’d gamwaith” (Jesus, or Oh Jesus! suffer not iniquity).

The first Eisteddfod of which any account seems to have descended to us was one held on the banks of the Conway in the 6th century, under the auspices of Maelgwn Gwynedd, prince of North Wales. Maelgwn on this occasion, in order to prove the superiority of vocal song over instrumental music, is recorded to have offered a reward to such bards and minstrels as should swim over the Conway. There were several competitors, but on their arrival on the opposite shore the harpers found themselves unable to play owing to the injury their harps had sustained from the water, while the bards were in as good tune as ever. King Cadwaladr also presided at an Eisteddfod about the middle of the 7th century.

Griffith ap Cynan, prince of North Wales, who had been born in Ireland, brought with him from that country many Irish musicians, who greatly improved the music of Wales. During his long reign of 56 years he offered great encouragement to bards, harpers and minstrels, and framed a code of laws for their better regulation. He held an Eisteddfod about the beginning of the 12th century at Caerwys in Flintshire, “to which there repaired all the musicians of Wales, and some also from England and Scotland.” For many years afterwards the Eisteddfod appears to have been held triennially, and to have enforced the rigid observance of the enactments of Griffith ap Cynan. The places at which it was generally held were Aberffraw, formerly the royal seat of the princes of North Wales; Dynevor, the royal castle of the princes of South Wales; and Mathrafal, the royal palace of the princes of Powys: and in later times Caerwys in Flintshire received that honourable distinction, it having been the princely residence of Llewelyn the Last. Some of these Eisteddfodau were conducted in a style of great magnificence, under the patronage of the native princes. At

Christmas 1107 Cadwgan, the son of Bleddyn ap Cynfyn, prince of Powys, held an Eisteddfod in Cardigan Castle, to which he invited the bards, harpers and minstrels, "the best to be found in all Wales"; and "he gave them chairs and subjects of emulation according to the custom of the feasts of King Arthur." In 1176 Rhys ab Gruffydd, prince of South Wales, held an Eisteddfod in the same castle on a scale of still greater magnificence, it having been proclaimed, we are told, a year before it took place, "over Wales, England, Scotland, Ireland and many other countries."

On the annexation of Wales to England, Edward I. deemed it politic to sanction the bardic Eisteddfod by his famous statute of Rhuddlan. In the reign of Edward III. Ifor Hael, a South Wales chieftain, held one at his mansion. Another was held in 1451, with the permission of the king, by Griffith ab Nicholas at Carmarthen, in princely style, where Dafydd ab Edmund, an eminent poet, signalized himself by his wonderful powers of versification in the Welsh metres, and whence "he carried home on his shoulders the silver chair" which he had fairly won. Several Eisteddfodau, were held, one at least by royal mandate, in the reign of Henry VII. In 1523 one was held at Caerwys before the chamberlain of North Wales and others, by virtue of a commission issued by Henry VIII. In the course of time, through relaxation of bardic discipline, the profession was assumed by unqualified persons, to the great detriment of the regular bards. Accordingly in 1567 Queen Elizabeth issued a commission for holding an Eisteddfod at Caerwys in the following year, which was duly held, when degrees were conferred on 55 candidates, including 20 harpers. From the terms of the royal proclamation we find that it was then customary to bestow "a silver harp" on the chief of the faculty of musicians, as it had been usual to reward the chief bard with "a silver chair." This was the last Eisteddfod appointed by royal commission, but several others of some importance were held during the 16th and 17th centuries, under the patronage of the earl of Pembroke, Sir Richard Neville, and other influential persons. Amongst these the last of any particular note was one held in Bewper Castle, Glamorgan, by Sir Richard Basset in 1681.

During the succeeding 130 years Welsh nationality was at its lowest ebb, and no general Eisteddfod on a large scale appears to have been held until 1819, though several small ones were held under the auspices of the Gwyneddigion Society, established in 1771,—the most important being those at Corwen (1789), St Asaph (1790) and Caerwys (1798).

At the close of the Napoleonic wars, however, there was a general revival of Welsh nationality, and numerous Welsh literary societies were established throughout Wales, and in the principal English towns. A large Eisteddfod was held under distinguished patronage at Carmarthen in 1819, and from that time to the present they have been held (together with numerous local Eisteddfodau), almost without intermission, annually. The Eisteddfod at Llangollen in 1858 is memorable for its archaic character, and the attempts then made to revive the ancient ceremonies, and restore the ancient vestments of druids, bards and ovates.

To constitute a provincial Eisteddfod it is necessary that it should be proclaimed by a graduated bard of a Gorsedd a year and a day before it takes place. A local one may be held without such a proclamation. A provincial Eisteddfod generally lasts three, sometimes four days. A president and a conductor are appointed for each day. The proceedings commence with a Gorsedd meeting, opened with sound of trumpet and other ceremonies, at which candidates come forward and receive bardic degrees after satisfying the presiding bard as to their fitness. At the subsequent meetings the president gives a brief address; the bards follow with poetical addresses; adjudications are made, and prizes and medals with suitable devices are given to the successful competitors for poetical, musical and prose compositions, for the best choral and solo singing, and singing with the harp or "Pennillion singing"[2] as it is called, for the best playing on the harp or stringed or wind instruments, as well as occasionally for the best specimens of handicraft and art. In the evening of each day a concert is given, generally attended by very large numbers. The great day of the Eisteddfod is the "chair" day—usually the third or last day—the grand event of the Eisteddfod being the adjudication on the chair subject, and the chairing and investiture of the fortunate winner. This is the highest object of a Welsh bard's ambition. The ceremony is an imposing one, and is performed with sound of trumpet. (See also the articles Bard, Celt: *Celtic Literature*, and Wales.)

(R. W.*)

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1 According to the Welsh Triads and other historical records, the *Gorsedd* or assembly (an essential part of the modern Eisteddfod, from which indeed the latter sprung) is as old at least as the time of Prydain the son of Ædd the Great, who lived many centuries before the Christian era. Upon the destruction of the political ascendancy of the Druids, the Gorsedd lost its political importance, though it seems to have long afterwards retained its institutional character as the medium for preserving the laws, doctrines and traditions of bardism.

2 According to Jones's *Bardic Remains*, "To sing 'Pennillion' with a Welsh harp is not so easily accomplished as may be imagined. The singer is obliged to follow the harper, who may change the tune, or perform variations *ad libitum*, whilst the vocalist must keep time, and end precisely with the strain. The singer does not commence with the harper, but takes the strain up at the second, third or fourth bar, as best suits the 'pennill' he intends to sing.... Those are considered the best singers who can adapt stanzas of various metres to one melody, and who are acquainted with the twenty-four measures according to the bardic laws and rules of composition."

EJECTMENT (Lat. e, out, and *jacere*, to throw), in English law, an action for the recovery of the possession of land, together with damages for the wrongful withholding thereof. In the old classifications of actions, as real or personal, this was known as a mixed action, because its object was twofold, viz. to recover both the realty and personal damages. It should be noted that the term "ejectment" applies in law to distinct classes of proceedings—ejectments as between rival claimants to land, and ejectments as between those who hold, or have held, the relation of landlord and tenant. Under the Rules of the Supreme Court, actions in England for the recovery of land are commenced and proceed in the same manner as ordinary actions. But the historical interest attaching to the action of ejectment is so great as to render some account of it necessary.

The form of the action as it prevailed in the English courts down to the Common Law Procedure Act 1852 was a series of fictions, among the most remarkable to be found in the entire body of English law. A, the person claiming title to land, delivered to B, the person in possession, a declaration in ejectment in which C and D, fictitious persons, were plaintiff and defendant. C stated that A had devised the land to him for a term of years, and that he had been ousted by D. A notice signed by D informed B of the proceedings, and advised him to apply to be made defendant in D's place, as he, D, having no title, did not intend to defend the suit. If B did not so apply, judgment was given against D, and possession of the lands was given to A. But if B did apply, the Court allowed him to defend the action only on condition that he admitted the three fictitious averments—the lease, the entry and the ouster—which, together with title, were the four things necessary to maintain an action of ejectment. This having been arranged the action proceeded, B being made defendant instead of D. The names used for the fictitious parties were John Doe, plaintiff, and Richard Roe, defendant, who was called "the casual ejector." The explanation of these mysterious fictions is this. The writ *de ejectione firmae* was invented about the beginning of the reign of Edward III. as a remedy to a lessee for years who had been ousted of his term. It was a writ of trespass, and carried damages, but in the time of Henry VII., if not before that date, the courts of common law added thereto a species of remedy neither warranted by the original writ nor demanded by the declaration, viz. a judgment to recover so much of the term as was still to run, and a writ of possession thereupon. The next step was to extend the remedy—limited originally to leaseholds—to cases of disputed title to freeholds. This was done indirectly by the claimant entering on the land and there making a lease for a term of years to another person; for it was only a term that could be recovered by the action, and to create a term required actual possession in the granter. The lessee remained on the land, and the next person who entered even by chance was accounted an ejector of the lessee, who then served upon him a writ of trespass and ejectment. The case then went to trial as on a common action of 138
trespass; and the claimant's title, being the real foundation of the lessee's right, was thus indirectly determined. These proceedings might take place without the knowledge of the person really in possession; and to prevent the abuse of the action a rule was laid down that the plaintiff in ejectment must give notice to the party in possession, who might then come in and defend the action. When the action came into general use as a mode of trying the title to freeholds, the actual entry, lease and ouster which were necessary to found the action were attended with much inconvenience, and accordingly Lord Chief Justice Rolle during the Protectorate (*c.* 1657) substituted for them the fictitious averments already described. The action of ejectment is now only a curiosity of legal history. Its fictitious suitors were swept away by the Common Law Procedure Act of 1852. A form of writ was prescribed, in which the person in possession of the disputed premises by name and all persons entitled to defend the possession were informed that the plaintiff claimed to be entitled to possession, and required to appear in court to defend the possession of the property or such part of it as they should think fit. In the form of the writ and in some other respects ejectment still differed from other actions. But, as already mentioned, it has now been assimilated (under the name of action for the recovery of lands) to ordinary actions by the Rules of the Supreme Court. It is commenced by writ of summons, and—subject to the rules as to summary judgments (*v. inf.*)—proceeds along the usual course of pleadings and trial to judgment; but is subject to one special rule, viz: that except by leave of the Court or a judge the only claims which may be joined with one for recovery of land are claims in respect of arrears of rent or double value for holding over, or mesne profits (*i.e.* the value of the land during the period of illegal possession), or damages for breach of a contract under which the premises are held or for any wrong or injury to the premises claimed (R.S.C., O. xviii. r. 2). These claims were formerly recoverable by an independent action.

With regard to actions for the recovery of land—apart from the relationship of landlord and tenant—the only point that need be noted is the presumption of law in favour of the actual possessor of the land in dispute. Where the action is brought by a landlord against his tenant, there is of course no presumption against the landlord's title arising from the tenant's possession. By the Common Law Procedure Act 1852 (ss. 210-212) special provision was made for the prompt recovery of demised premises where half a year's rent was in arrear and the landlord was entitled to re-enter for non-payment. These provisions are still in force, but advantage is now more generally taken of the summary judgment procedure introduced by the Rules of the Supreme Court (Order 3, r. 6.). This procedure may be adopted when (*a*) the tenant's term has expired, (*b*) or has been duly determined by notice to quit, or (*c*) has become liable to forfeiture for non-payment of rent, and applies not only to the tenant but to persons claiming under him. The writ is specially endorsed with the plaintiff's claim to recover the land with or without rent or mesne profits, and summary judgment obtained if no substantial defence is disclosed. Where an action to recover land is brought against the tenant by a person claiming adversely to the landlord, the tenant is bound, under penalty of forfeiting the value of three years' improved or rack rent of the premises, to give notice to the landlord in order that he may appear and defend his title. Actions for the recovery of land, other than land belonging to spiritual corporations and to the crown, are barred in 12 years (Real Property Limitation Acts 1833 (s. 29) and 1874 (s. 1). A landlord can recover possession in the county court (i.) by an action for the recovery of possession, where neither the value of the premises nor the rent exceeds £100 a year, and the tenant is holding over (County Courts Acts of 1888, s. 138, and 1903, s. 3); (ii.) by "an action of ejectment," where (*a*) the value or rent of the premises does not exceed £100, (*b*) half a year's rent is in arrear, and (*c*) no sufficient distress (see Rent) is to be found on the premises (Act of 1888, s. 139; Act of 1903, s. 3; County Court Rules 1903, Ord. v. rule 3). Where a tenant at a rent not exceeding £20 a year of premises at will, or for a term not exceeding 7 years, refuses nor neglects, on the determination or expiration of his interest, to deliver up possession, such possession may be recovered by proceedings before justices under the Small Tenements Recovery Act 1838, an enactment which has been extended to the recovery of allotments. Under the Distress for Rent Act 1737, and the Deserted Tenements Act 1817, a landlord can have himself put by the order of two justices into premises deserted by the tenant where half a year's rent is owing and no sufficient distress can be found.

In *Ireland*, the practice with regard to the recovery of land is regulated by the Rules of the Supreme Court 1891, made under the Judicature (Ireland) Act 1877; and resembles that of England. Possession may be recovered summarily by a special indorsement of the writ, as in England; and there are analogous provisions with regard to the recovery of small tenements (see Land Act, 1860 ss. 84 and 89). The law with regard to the ejectment or eviction of tenants is consolidated by the Land Act 1860. (See ss. 52-66, 68-71, and further under Landlord and Tenant.)

In *Scotland*, the recovery of land is effected by an action of "removing" or summary ejection. In the case of a tenant "warning" is necessary unless he is bound by his lease to remove without warning. In the case of possessors without title, or a title merely precarious, no warning is needed. A summary process of removing from small holdings is provided for by Sheriff Courts (Scotland) Acts of 1838 and 1851.

In the United States, the old English action of ejectment was adopted to a very limited extent, and where it was so adopted has often been superseded, as in Connecticut, by a single action for all cases of ouster, disseisin or ejectment. In this action, known as an action of disseisin or ejectment, both possession of the land and damages may be recovered. In some of the states a tenant against whom an action of ejectment is brought by a stranger is bound under a penalty, as in England, to give notice of the claim to the landlord in order that he may appear and defend his title.

In *French law* the landlord's claim for rent is fairly secured by the hypothec, and by summary powers which exist for the seizure of the effects of defaulting tenants. Eviction or annulment of a lease can only be obtained through the judicial tribunals. The Civil Code deals with the position of a tenant in case of the sale of the property leased. If the lease is by authentic act (*acte authentique*) or has an ascertained date, the purchaser cannot evict the tenant unless a right to do so was reserved on the lease (art. 1743), and then only on payment of an indemnity (arts. 1744-1747). If the lease is not by authentic act, or has not an ascertained date, the purchaser is not liable for indemnity (art. 1750). The tenant of rural lands is bound to give the landlord notice of acts of usurpation (art. 1768). There are analogous provisions in the Civil Codes of Belgium (arts. 1743 et seq.), Holland (arts. 1613, 1614), Portugal (art. 1572); and see the German Civil Code (arts. 535 et seq.). In many of the colonies there are statutory provisions for the recovery of land or premises on the lines of English law (cf. Ontario, Rev. Stats. 1897, c. 170. ss. 19 et seq.; Manitoba, Rev. Stats. 1902, *c.* 1903). In others (*e.g.* New Zealand, Act. No. 55 of 1893, ss. 175-187; British Columbia, Revised Statutes, 1897, c. 182: Cyprus, Ord. 15 of 1895) there has been legislation similar to the Small Tenements Recovery Act 1838.

Authorities.—*English Law*: Cole on *Ejectment*; Digby, *History of Real Property* (3rd ed., London, 1884); Pollock and Maitland, *History of English Law* (Cambridge, 1895); Foa, *Landlord and Tenant* (4th ed., London, 1907); Fawcett, *Landlord and Tenant* (London, 1905). *Irish Law*: Nolan and Kane's *Statutes relating to the Law of Landlord and Tenant* (5th ed., Dublin, 1898); Wylie's *Judicature Acts* (Dublin, 1900). *Scots Law*: Hunter on *Landlord and Tenant* (4th ed., Edin., 1878); Erskine's *Principles* (20th ed., Edin., 1903). *American Law: Two Centuries' Growth of American Law* (New York and London, 1901); Bouvier's *Law Dictionary* (Boston and London, 1897); Stimson, *American Statute Law* (Boston, 1886).

(A. W. R.)

EKATERINBURG, a town of Russia, in the government of Perm, 311 m. by rail S.E. of the town of Perm, on the Iset river, near the E. foot of the Ural Mountains, in 56° 49′ N. and 60° 35′ E., at an altitude of 870 ft. 139
above sea-level. It is the most important town of the Urals. Pop. (1860) 19,830; (1897) 55,488. The streets are broad and regular, and several of the houses of palatial proportions. In 1834 Ekaterinburg was made the see of a suffragan bishop of the Orthodox Greek Church. There are two cathedrals—St Catherine's, founded in 1758, and that of the Epiphany, in 1774—and a museum of natural history, opened in 1853. Ekaterinburg is the seat of the central mining administration of the Ural region, and has a chemical laboratory for the assay of gold, a mining school, the Ural Society of Naturalists, and a magnetic and meteorological observatory. Besides the government mint for copper coinage, which dates from 1735, the government engineering works, and the imperial factory for the cutting and polishing of malachite, jasper, marble, porphyry and other ornamental stones, the industrial establishments comprise candle, paper, soap and machinery works, flour and woollen mills, and tanneries. There is a lively trade in cattle, cereals, iron, woollen and silk goods, and colonial products; and two important fairs are held annually. Nearly forty gold and platinum mines, over thirty iron-works, and numerous other factories are scattered over the district, while wheels, travelling boxes, hardware, boots and so forth are extensively made in the villages. Ekaterinburg took its origin from the mining establishments founded by Peter the Great in 1721, and received its name in honour of his wife, Catherine I. Its development was greatly promoted in 1763 by the diversion of the Siberian highway from Verkhoturye to this place.

EKATERINODAR, a town of South Russia, chief town of the province of Kubañ, on the right bank of the river Kubañ, 85 m. E.N.E. of Novo-rossiysk on the railway to Rostov-on-Don, and in 45° 3′ N. and 38° 50′ E. It is badly built, on a swampy site exposed to the inundations of the river; and its houses, with few exceptions, are slight structures of wood and plaster. Founded by Catherine II. in 1794 on the site of an old town called Tmutarakan, as a small fort and Cossack settlement, its population grew from 9620 in 1860 to 65,697 in 1897. It has various technical schools, an experimental fruit-farm, a military hospital, and a natural history museum. A considerable trade is carried on, especially in cereals.

EKATERINOSLAV, a government of south Russia, having the governments of Poltava and Kharkov on the N., the territory of the Don Cossacks on the E., the Sea of Azov and Taurida on the S., and Kherson on the W. Area, 24,478 sq. m. Its surface is undulating steppe, sloping gently south and north, with a few hills reaching 1200 ft. in the N.E., where a slight swelling (the Don Hills) compels the Don to make a great curve eastwards. Another chain of hills, to which the eastward bend of the Dnieper is due, rises in the west. These hills have a crystalline core (granites, syenites and diorites), while the surface strata belong to the Carboniferous, Permian, Cretaceous and Tertiary formations. The government is rich in minerals, especially in coal—the mines lie in the middle of the Donets coalfield—iron ores, fireclay and rock-salt, and every year the mining output increases in quantity, especially of coal and iron. Granite, limestone, grindstone, slate, with graphite, manganese and mercury are found. The government is drained by the Dnieper, the Don and their tributaries (*e.g.* the Donets and Volchya) and by several affluents (*e.g.* the Kalmius) of the Sea of Azov. The soil is the fertile black earth, but the crops occasionally suffer from drought, the average annual rainfall being only 15 in. Forests are scarce. Pop. (1860) 1,138,750; (1897) 2,118,946, chiefly Little Russians, with Great Russians, Greeks (48,740), Germans (80,979), Rumanians and a few gypsies. Jews constitute 4.7% of the population. The estimated population in 1906 was 2,708,700.

Wheat and other cereals are extensively grown; other noteworthy crops are potatoes, tobacco and grapes. Nearly 40,000 persons find occupation in factories, the most important being iron-works and agricultural machinery works, though there are also tobacco, glass, soap and candle factories, potteries, tanneries and breweries. In the districts of Mariupol the making of agricultural implements and machinery is carried on extensively as a domestic industry in the villages. Bees are kept in very considerable numbers. Fishing employs many persons in the Don and the Dnieper. Cereals are exported in large quantities via the Dnieper, the Sevastopol railway, and the port of Mariupol. The chief towns of the eight districts, with their populations in 1897, are Ekaterinoslav (135,552 inhabitants in 1900), Alexandrovsk (28,434), Bakhmut (30,585), Mariupol (31,772), Novomoskovsk (12,862), Pavlograd (17,188), Slavyanoserbsk (3120), and Verkhne-dnyeprovsk (11,607).

EKATERINOSLAV, a town of Russia, capital of the government of the same name, on the right bank of the Dnieper above the rapids, 673 m. by rail S.S.W. of Moscow, in 48° 21′ N. and 35° 4′ E., at an altitude of 210 ft. Pop. (1861) 18,881, without suburbs; (1900) 135,552. If the suburb of Novyikoindak be included, the town extends for upwards of 4 m. along the river. The oldest part lies very low and is much exposed to floods. Contiguous to the towns on the N.W. is the royal village of Novyimaidani or the New Factories. The bishop’s palace, mining academy, archaeological museum and library are the principal public buildings. The house now occupied by the Nobles Club was formerly inhabited by the author and statesman Potemkin. Ekaterinoslav is a rapidly growing city, with a number of technical schools, and is an important depot for timber floated down the Dnieper, and also for cereals. Its iron-works, flour-mills and agricultural machinery works give occupation to over 5000 persons. In fact since 1895 the city has become the centre of numerous Franco-Belgian industrial undertakings. In addition to the branches just mentioned, there are tobacco factories and breweries. Considerable trade is carried on in cattle, cereals, horses and wool, there being three annual fairs. On the site of the city there formerly stood the Polish castle of Koindak, built in 1635, and destroyed by the Cossacks. The existing city was founded by Potemkin in 1786, and in the following year Catherine II. laid the foundation-stone of the cathedral, though it was not actually built until 1830-1835. On the south side of it is a bronze statue of the empress, put up in 1846. Paul I. changed the name of the city to Novo-rossiysk, but the original name was restored in 1802.

EKHOF, KONRAD (1720-1778), German actor, was born in Hamburg on the 12th of August 1720. In 1739 he became a member of Johann Friedrich Schönemann’s (1704-1782) company in Lüneburg, and made his first appearance there on the 15th of January 1740 as Xiphares in Racine’s *Mithridate*. From 1751 the Schönemann company performed mainly in Hamburg and at Schwerin, where Duke Christian Louis II. of Mecklenburg-Schwerin made them comedians to the court. During this period Ekhof founded a theatrical academy, which, though short-lived, was of great importance in helping to raise the standard of German acting and the status of German actors. In 1757 Ekhof left Schönemann to join Franz Schuch’s company at Danzig; but he soon returned to Hamburg, where, in conjunction with two other actors, he succeeded Schönemann in the direction of the company. He resigned this position, however, in favour of H.G. Koch, with whom he acted until 1764, when he joined K.E. Ackermann’s company. In 1767 was founded the National Theatre at Hamburg, made famous by Lessing’s *Hamburgische Dramaturgie*, and Ekhof was the leading member of the company. After the failure of the enterprise Ekhof was for a time in Weimar, and ultimately became co-director of the new court theatre at Gotha. This, the first permanently established theatre in Germany, was opened on the 2nd of October 1775. Ekhof’s reputation was now at its height; Goethe called him the only German tragic actor; and in 1777 he acted with Goethe and Duke Charles Augustus at a private performance at Weimar, dining afterwards with the poet at the ducal table. He died on the 16th of June 1778. His versatility may be judged from the fact that in the comedies of Goldoni and Molière he was no less successful than in the tragedies of Lessing and Shakespeare. He was regarded by his contemporaries as an unsurpassed exponent of naturalness on the stage; and in this respect he has been not unfairly compared with Garrick. His fame, however, was rapidly eclipsed by that of Friedrich U.L. Schröder. His literary efforts were chiefly confined to translations from French authors.

See H. Uhde, biography of Ekhof in vol. iv. of *Der neue Plutarch* (1876), and J. Rüschner, *K. Ekhofs Leben und Wirken* (1872). Also H. Devrient, *J.F. Schönemann und seine Schauspielergesellschaft* (1895).

EKRON (better, as in the Septuagint and Josephus, Accaron, Ἀκκαρών), a royal city of the Philistines commonly identified with the modern Syrian village of ‘Aḳir, 5 m. from Ramleh, on the southern slope of a low ridge separating the plain of Philistia from Sharon. It lay inland and off the main line of traffic. Though included by the Israelites within the limits of the tribe of Judah, and mentioned in Judges xix. as one of the cities of Dan, it was in Philistine possession in the days of Samuel, and apparently maintained its independence. According to the narrative of the Hebrew text, here differing from the Greek text and Josephus (which read Askelon), it was the last town to which the ark was transferred before its restoration to the Israelites. Its maintenance of a sanctuary of Baal Zebub is mentioned in 2 Kings i. From Assyrian inscriptions it has been gathered that Padi, king of Ekron, was for a time the vassal of Hezekiah of Judah, but regained his independence when the latter was hard pressed by Sennacherib. A notice of its history in 147 B.C. is found in 1 Macc. x. 89; after the fall of Jerusalem A.D. 70 it was settled by Jews. At the time of the crusades it was still a large village. Recently a Jewish agricultural colony has been settled there. The houses are built of mud, and in the absence of visible remains of antiquity, the identification of the site is questionable. The neighbourhood is fertile.

(R. A. S. M.)

ELABUGA, a town of Russia, in the government of Vyatka, on the Kama river, 201 m. by steamboat down the Volga from Kazan and then up the Kama. It has flour-mills, and carries on a brisk trade in exporting corn. Pop. (1897) 9776.

The famous *Ananiynskiy Mogilnik* (burial-place) is on the right bank of the Kama, 3 m. above the town. It was discovered in 1858, was excavated by Alabin, Lerch and Nevostruyev, and has since supplied extremely valuable collections belonging to the Stone, Bronze and Iron Ages. It consisted of a mound, about 500 ft. in circumference, adorned with decorated stones (which have disappeared), and contained an inner wall, 65 ft. in circumference, made of uncemented stone flags. Nearly fifty skeletons were discovered, mostly lying upon charred logs, surrounded with cinerary urns filled with partially burned bones. A great variety of bronze decorations and glazed clay pearls were strewn round the skeletons. The knives, daggers and arrowpoints are of slate, bronze and iron, the last two being very rough imitations of stone implements. One of the flags bore the image of a man, without moustaches or beard, dressed in a costume and helmet recalling those of the Circassians.

ELAM, the name given in the Bible to the province of Persia called Susiana by the classical geographers, from Susa or Shushan its capital. In one passage, however (Ezra iv. 9), it is confined to Elymais, the north-western part of the province, and its inhabitants distinguished from those of Shushan, which elsewhere (Dan. viii. 2) is placed in Elam. Strabo (xv. 3. 12, &c.) makes Susiana a part of Persia proper, but a comparison of his account with those of Ptolemy (vi. 3. 1, &c.) and other writers would limit it to the mountainous district to the east of Babylonia, lying between the Oroatis and the Tigris, and stretching from India to the Persian Gulf. Along with this mountainous district went a fertile low tract of country on the western side, which also included the marshes at the mouths of the Euphrates and Tigris and the north-eastern coast land of the Gulf. This low tract, though producing large quantities of grain, was intensely hot in summer; the high regions, however, were cool and well watered.

The whole country was occupied by a variety of tribes, speaking agglutinative dialects for the most part, though the western districts were occupied by Semites. Strabo (xi. 13. 3, 6), quoting from Nearchus, seems to include the Susians under the Elymaeans, whom he associates with the Uxii, and places on the frontiers of Persia and Susa; but Pliny more correctly makes the Eulaeus the boundary between Susiana and Elymais (*N.H.* vi. 29-31). The Uxii are described as a robber tribe in the mountains adjacent to Media, and their name is apparently to be identified with the title given to the whole of Susiana in the Persian cuneiform inscriptions, *Uwaja*, *i.e.* “Aborigines.” Uwaja is probably the origin of the modern Khuzistan, though Mordtmann would derive the latter from خوز “a sugar-reed.” Immediately bordering on the Persians were the Amardians or Mardians, as well as the people of Khapirti (Khatamti, according to Scheil), the name given to Susiana in the Neo-Susian texts. Khapirti appears as Apir in the inscriptions of Mal-Amir, which fix the locality of the district. Passing over the Messabatae, who inhabited a valley which may perhaps be the modern Māh-Sabadan, as well as the level district of Yamutbal or Yatbur which separated Elam from Babylonia, and the smaller districts of Characene, Cabandene, Corbiana and Gabiene mentioned by classical authors, we come to the fourth principal tribe of Susiana, the Cissii (Aesch. *Pers.* 16; Strabo xv. 3. 2) or Cossaei (Strabo xi. 5. 6, xvi. 11. 17; Arr. *Ind.* 40; Polyb. v. 54, &c.), the Kassi of the cuneiform inscriptions. So important were they, that the whole of Susiana was sometimes called Cissia after them, as by Herodotus (iii. 91, v. 49, &c.). In fact Susiana was only a late name for the country, dating from the time when Susa had been made a capital of the Persian empire. In the Sumerian texts of Babylonia it was called Numma, “the Highlands,” of which Elamtu or Elamu, “Elam,” was the Semitic translation. Apart from Susa, the most important part of the country was Anzan (Anshan, contracted Assan), where the native population maintained itself unaffected by Semitic intrusion. The exact position of Anzan is still disputed, but it probably included originally the site of Susa and was distinguished from it only when Susa became the seat of a Semitic government. In the lexical tablets Anzan is given as the equivalent of Elamtu, and the native kings entitle themselves kings of “Anzan and Susa,” as well as “princes of the Khapirti.”

The principal mountains of Elam were on the north, called Charbanus and Cambalidus by Pliny (vi. 27, 31), and belonging to the Parachoathras chain. There were numerous rivers flowing into either the Tigris or the Persian Gulf. The most important were the Ulai or Eulaeus (*Kūran*) with its tributary the Pasitigris, the Choaspes (*Kerkhah*), the Coprates (river of *Diz* called Itītē in the inscriptions), the Hedyphon or Hedypnus (*Jerrāhi*), and the Croatis (*Hindyan*), besides the monumental Surappi and Ukni, perhaps to be identified with the Hedyphon and Oroatis, which fell into the sea in the marshy region at the mouth of the Tigris. Shushan or Susa, the capital now marked by the mounds of *Shush*, stood near the junction of the Choaspes and Eulaeus (see Susa); and Badaca, Madaktu in the inscriptions, lay between the *Shapur* and the river of *Diz*. Among the other chief cities mentioned in the inscriptions may be named Naditu, Khaltemas, Din-sar, Bubilu, Bit-imbi, Khidalu and Nagitu on the sea-coast. Here, in fact, lay some of the oldest and wealthiest towns, the sites of which have, however, been removed inland by the silting up of the shore. J. de Morgan’s excavations at Susa have thrown a flood of light on the early history of Elam and its relations to Babylon. The earliest settlement there goes back to neolithic times, but it was already a fortified city when Elam was conquered by Sargon of Akkad (3800 B.C.) and Susa became the seat of a Babylonian viceroy. From this time onward for many centuries it continued under Semitic suzerainty, its high-priests, also called “Chief Envoys of Elam, Sippara and Susa,” bearing sometimes Semitic, sometimes native “Anzanite” names. One of the kings of the dynasty of Ur built at Susa. Before the rise of the First Dynasty of Babylon, however, Elam had recovered its independence, and in 2280 B.C. the Elamite king Kutur-Nakhkhunte made a raid in Babylonia and carried away from Erech the image of the goddess Nanā. The monuments of many of his successors have been discovered by de Morgan and their inscriptions deciphered by v. Scheil. One of them was defeated by Ammi-zadoq of Babylonia (*c.* 2100 B.C.); another would have been the Chedor-laomer (Kutur-Lagamar) of Genesis 141
xiv. One of the greatest builders among them was Untas-Gal (the pronunciation of the second element in the name is uncertain). About 1330 B.C. Khurba-tila was captured by Kuri-galzu III., the Kassite king of Babylonia, but a later prince Kidin-Khutrutas avenged his defeat, and Sutruk-Nakhkhunte (1220 B.C.) carried fire and sword through Babylonia, slew its king Zamama-sum-iddin and carried away a stela of Naram-Sin and the famous code of laws of Khammurabi from Sippara, as well as a stela of Manistusu from Akkuttum or Akkad. He also conquered the land of Asnunnak and carried off from Padan a stela belonging to a refugee from Malatia. He was succeeded by his son who was followed on the throne by his brother, one of the great builders of Elam. In 750 B.C. Umbadara was king of Elam; Khumban-igas was his successor in 742 B.C. In 720 B.C. the latter prince met the Assyrians under Sargon at Dur-ili in Yamutbal, and though Sargon claims a victory the result was that Babylonia recovered its independence under Merodach-baladan and the Assyrian forces were driven north. From this time forward it was against Assyria instead of Babylonia that Elam found itself compelled to exert its strength, and Elamite policy was directed towards fomenting revolt in Babylonia and assisting the Babylonians in their struggle with Assyria. In 716 B.C. Khumban-igas died and was followed by his nephew, Sutruk-Nakhkhunte. He failed to make head against the Assyrians; the frontier cities were taken by Sargon and Merodach-baladan was left to his fate. A few years later (704 B.C.) the combined forces of Elam and Babylonia were overthrown at Kis, and in the following year the Kassites were reduced to subjection. The Elamite king was dethroned and imprisoned in 700 B.C. by his brother Khallusu, who six years later marched into Babylonia, captured the son of Sennacherib, whom his father had placed there as king, and raised a nominee of his own, Nergal-yusezib, to the throne. Khallusu was murdered in 694 B.C., after seeing the maritime part of his dominions invaded by the Assyrians. His successor Kudur-Nakhkhunte invaded Babylonia; he was repulsed, however, by Sennacherib, 34 of his cities were destroyed, and he himself fled from Madaktu to Khidalu. The result was a revolt in which he was killed after a reign of ten months. His brother Umman-menan at once collected allies and prepared for resistance to the Assyrians. But the terrible defeat at Khalulē broke his power; he was attacked by paralysis shortly afterwards, and Khumba-Khaldas II. followed him on the throne (689 B.C.). The new king endeavoured to gain Assyrian favour by putting to death the son of Merodach-baladan, but was himself murdered by his brothers Urtaki and Teumman (681 B.C.), the first of whom seized the crown. On his death Teumman succeeded and almost immediately provoked a quarrel with Assur-bani-pal by demanding the surrender of his nephews who had taken refuge at the Assyrian court. The Assyrians pursued the Elamite army to Susa, where a battle was fought on the banks of the Eulaeus, in which the Elamites were defeated, Teumman captured and slain, and Umman-igas, the son of Urtaki, made king, his younger brother Tammaritu being given the district of Khidalu. Umman-igas afterwards assisted in the revolt of Babylonia under Samas-sum-yukin, but his nephew, a second Tammaritu, raised a rebellion against him, defeated him in battle, cut off his head and seized the crown. Tammaritu marched to Babylonia; while there, his officer Inda-bigas made himself master of Susa and drove Tammaritu to the coast whence he fled to Assur-bani-pal. Inda-bigas was himself overthrown and slain by a new pretender, Khumba-Khaldas III., who was opposed, however, by three other rivals, two of whom maintained themselves in the mountains until the Assyrian conquest of the country, when Tammaritu was first restored and then imprisoned, Elam being utterly devastated. The return of Khumba-Khaldas led to a fresh Assyrian invasion; the Elamite king fled from Madaktu to Dur-undasi; Susa and other cities were taken, and the Elamite army almost exterminated on the banks of the Itītē. The whole country was reduced to a desert, Susa was plundered and razed to the ground, the royal sepulchres were desecrated, and the images of the gods and of 32 kings “in silver, gold, bronze and alabaster,” were carried away. All this must have happened about 640 B.C. After the fall of the Assyrian empire Elam was occupied by the Persian Teispes, the forefather of Cyrus, who, accordingly, like his immediate successors, is called in the inscriptions “king of Anzan.” Susa once more became a capital, and on the establishment of the Persian empire remained one of the three seats of government, its language, the Neo-Susian, ranking with the Persian of Persepolis and the Semitic of Babylon as an official tongue. In the reign of Darius, however, the Susianians attempted to revolt, first under Assina or Atrina, the son of Umbadara, and later under Martiya, the son of Issainsakria, who

called himself Immanes; but they gradually became completely Aryanized, and their agglutinative dialects were supplanted by the Aryan Persian from the south-east.

Elam, “the land of the cedar-forest,” with its enchanted trees, figured largely in Babylonian mythology, and one of the adventures of the hero Gilgamesh was the destruction of the tyrant Khumbaba who dwelt in the midst of it. A list of the Elamite deities is given by Assur-bani-pal; at the head of them was In-Susinak, “the lord of the Susians,”—a title which went back to the age of Babylonian suzerainty,—whose image and oracle were hidden from the eyes of the profane. Nakhkhunte, according to Scheil, was the Sun-goddess, and Lagamar, whose name enters into that of Chedor-laomer, was borrowed from Semitic Babylonia.

See W.K. Loftus, *Chaldaea and Susiana* (1857); A. Billerbeck, *Susa* (1893); J. de Morgan, *Mémoires de la Délégation en Perse* (9 vols., 1899-1906).

(A. H. S.)

ELAND (= elk), the Dutch name for the largest of the South African antelopes (*Taurotragus oryx*), a species near akin to the kudu, but with horns present in both sexes, and their spiral much closer, being in fact screw-like instead of corkscrew-like. There is also a large dewlap, while old bulls have a thick forelock. In the typical southern form the body-colour is wholly pale fawn, but north of the Orange river the body is marked by narrow vertical white lines, this race being known as *T. oryx livingstonei*. In Senegambia the genus is represented by *T. derbianus*, a much larger animal, with a dark neck; while in the Bahr-el-Ghazal district there is a gigantic local race of this species (*T. derbianus giganteus*).

(R. L.*)

ELASTICITY. 1. Elasticity is the property of recovery of an original size or shape. A body of which the size, or shape, or both size and shape, have been altered by the application of forces may, and generally does, tend to return to its previous size and shape when the forces cease to act. Bodies which exhibit this tendency are said to be *elastic* (from Greek, ἐλαύνειν, to drive). All bodies are more or less elastic as regards size; and all solid bodies are more or less elastic as regards shape. For example: gas contained in a vessel, which is closed by a piston, can be compressed by additional pressure applied to the piston; but, when the additional pressure is removed, the gas expands and drives the piston outwards. For a second example: a steel bar hanging vertically, and loaded with one ton for each square inch of its sectional area, will have its length increased by about seven one-hundred-thousandths of itself, and its sectional area diminished by about half as much; and it will spring back to its original length and sectional area when the load is gradually removed. Such changes of size and shape in bodies subjected to forces, and the recovery of the original size and shape when the forces cease to act, become conspicuous when the bodies have the forms of thin wires or planks; and these properties of bodies in such forms are utilized in the construction of spring balances, carriage springs, buffers and so on.

It is a familiar fact that the hair-spring of a watch can be coiled and uncoiled millions of times a year for several years without losing its elasticity; yet the same spring can have its shape permanently altered by forces which are much greater than those to which it is subjected in the motion of the watch. The incompleteness of the recovery from the effects of great forces is as important a fact as the practical completeness of the recovery from the effects of comparatively small forces. The fact is referred to in the 142
distinction between “perfect” and “imperfect” elasticity; and the limitation which must be imposed upon the forces in order that the elasticity may be perfect leads to the investigation of “limits of elasticity” (see §§ 31, 32 below). Steel pianoforte wire is perfectly elastic within rather wide limits, glass within rather narrow limits; building stone, cement and cast iron appear not to be perfectly elastic within any limits, however narrow. When the limits of elasticity are not exceeded no injury is done to a material or structure by the action of the forces. The strength or weakness of a material, and the safety or insecurity of a structure, are thus closely related to the elasticity of the material and to the change of size or shape of the structure when subjected to forces. The “science of elasticity” is occupied with the more abstract side of this relation, viz. with the effects that are produced in a body of definite size, shape and constitution by definite forces; the “science of the strength of materials” is occupied with the more concrete side, viz. with the application of the results obtained in the science of elasticity to practical questions of strength and safety (see Strength of Materials).

2. *Stress.*—Every body that we know anything about is always under the action of forces. Every body upon which we can experiment is subject to the force of gravity, and must, for the purpose of experiment, be supported by other forces. Such forces are usually applied by way of pressure upon a portion of the surface of the body; and such pressure is exerted by another body in contact with the first. The supported body exerts an equal and opposite pressure upon the supporting body across the portion of surface which is common to the two. The same thing is true of two portions of the same body. If, for example, we consider the two portions into which a body is divided by a (geometrical) horizontal plane, we conclude that the lower portion supports the upper portion by pressure across the plane, and the upper portion presses downwards upon the lower portion with an equal pressure. The pressure is still exerted when the plane is not horizontal, and its direction may be obliquely inclined to, or tangential to, the plane. A more precise meaning is given to “pressure” below. It is important to distinguish between the two classes of forces: forces such as the force of gravity, which act all through a body, and forces such as pressure applied over a surface. The former are

named “body forces” or “volume forces,” and the latter “surface tractions.” The action between two portions of a body separated by a geometrical surface is of the nature of surface traction. Body forces are ultimately, when the volumes upon which they act are small enough, proportional to the volumes; surface tractions, on the other hand, are ultimately, when the surfaces across which they act are small enough, proportional to these surfaces. Surface tractions are always exerted by one body upon another, or by one part of a body upon another part, across a surface of contact; and a surface traction is always to be regarded as one aspect of a “stress,” that is to say of a pair of equal and opposite forces; for an equal traction is always exerted by the second body, or part, upon the first across the surface.

3. The proper method of estimating and specifying stress is a matter of importance, and its character is necessarily mathematical. The magnitudes of the surface tractions which compose a stress are estimated as so much force (in dynes or tons) per unit of area (per sq. cm. or per sq. in.). The traction across an assigned plane at an assigned point is measured by the mathematical limit of the fraction F/S, where F denotes the numerical measure of the force exerted across a small portion of the plane containing the point, and S denotes the numerical measure of the area of this portion, and the limit is taken by diminishing S indefinitely. The traction may act as “tension,” as it does in the case of a horizontal section of a bar supported at its upper end and hanging vertically, or as “pressure,” as it does in the case of a horizontal section of a block resting on a horizontal plane, or again it may act obliquely or even tangentially to the separating plane. Normal tractions are reckoned as positive when they are tensions, negative when they are pressures. Tangential tractions are often called “shears” (see § 7 below). Oblique tractions can always be resolved, by the vector law, into normal and tangential tractions. In a fluid at rest the traction across any plane at any point is normal to the plane, and acts as pressure. For the complete specification of the “state of stress” at any point of a body, we should require to know the normal and tangential components of the traction across every plane drawn through the point. Fortunately this requirement can be very much simplified (see §§ 6, 7 below).

4. In general let ν denote the direction of the normal drawn in a specified sense to a plane drawn through a point O of a body; and let T_ν denote the traction exerted across the plane, at the point O, by the portion of the body towards which ν is drawn upon the remaining portion. Then T_ν is a vector quantity, which has a definite magnitude (estimated as above by the limit of a fraction of the form F/S) and a definite direction. It can be specified completely by its components X_ν, Y_ν, Z_ν, referred to fixed rectangular axes of x, y, z. When the direction of ν is that of the axis of x, in the positive sense, the components are denoted by X_x, Y_x, Z_x; and a similar notation is used when the direction of ν is that of y or z, the suffix x being replaced by y or z.

5. Every body about which we know anything is always in a state of stress, that is to say there are always internal forces acting between the parts of the body, and these forces are exerted as surface tractions across geometrical surfaces drawn in the body. The body, and each part of the body, moves under the action of all the forces (body forces and surface tractions) which are exerted upon it; or remains at rest if these forces are in equilibrium. This result is expressed analytically by means of certain equations—the “equations of motion” or “equations of equilibrium” of the body.

Let ρ denote the density of the body at any point, X, Y, Z, the components parallel to the axes of x, y, z of the body forces, estimated as so much force per unit of mass; further let f_x, f_y, f_z denote the components, parallel to the same axes, of the acceleration of the particle which is momentarily at the point (x, y, z). The equations of motion express the result that the rates of change of the momentum, and of the moment of momentum, of any portion of the body are those due to the action of all the forces exerted upon the portion by other bodies, or by other portions of the same body. For the changes of momentum, we have three equations of the type

$$\int\int\int \rho\, X dx\, dy\, dz + \int\int X_\nu dS = \int\int\int \rho\, f_x dx\, dy\, dz, \tag{1}$$

in which the volume integrations are taken through the volume of the portion of the body, the surface integration is taken over its surface, and the notation X_ν is that of § 4, the direction of ν being that of the normal to this surface drawn outwards. For the changes of moment of momentum, we have three equations of the type

$$\int\int\int \rho\, (yZ - zY)\, dx\, dy\, dz + \int\int (yZ_\nu - zY_\nu)\, dS = \int\int\int \rho\, (yf_z - zf_y)\, dx\, dy\, dz. \tag{2}$$

The equations (1) and (2) are the equations of motion of any kind of body. The equations of equilibrium are obtained by replacing the right-hand members of these equations by zero.

6. These equations can be used to obtain relations between the values of X_ν, Y_ν, ... for different directions ν. When the equations are applied to a very small volume, it appears that the terms expressed by surface integrals would, unless they tend to zero limits in a higher order than the areas of the surfaces, be very great compared with the terms expressed by volume integrals. We conclude that the surface tractions on the portion of the body which is bounded by any very small closed surface, are ultimately in equilibrium. When this result is interpreted for a small portion in the shape of a tetrahedron, having three of its faces at right angles to the co-ordinate axes, it leads to three equations of the type

$$X_\nu = X_x \cos(x, \nu) + X_y \cos(y, \nu) + X_z \cos(z, \nu), \tag{1}$$

where ν is the direction of the normal (drawn outwards) to the remaining face of the tetrahedron, and (x, ν) ... denote the angles which this normal makes with the axes. Hence X_ν, ... for any direction ν are expressed in terms of X_x,.... When the above result is interpreted for a very small portion in the shape of a cube, having its edges parallel to the co-ordinate axes, it leads to the equations

$$Y_z = Z_y, \qquad Z_x = X_z, \qquad X_y = Y_x. \tag{2}$$

When we substitute in the general equations the particular results which are thus obtained, we find that the equations of motion take such forms as

$$\partial X_x \qquad \partial X_y \qquad \partial Z_x$$

$$\rho X + \frac{}{\partial x} + \frac{}{\partial y} + \frac{}{\partial z} = \rho f_x, \quad (3)$$

and the equations of moments are satisfied identically. The equations of equilibrium are obtained by replacing the right-hand members by zero.

7. A state of stress in which the traction across any plane of a set of parallel planes is normal to the plane, and that across any perpendicular plane vanishes, is described as a state of “simple tension” (“simple pressure” if the traction is negative). A state of stress in which the traction across any plane is normal to the plane, and the traction is the same for all planes passing through any point, is described as a state of “uniform tension” (“uniform pressure” if the traction is negative). Sometimes the phrases “isotropic tension” and “hydrostatic pressure” are used instead of “uniform” tension or pressure. The distinction between the two states, simple tension and uniform tension, is illustrated in fig. 1. 143

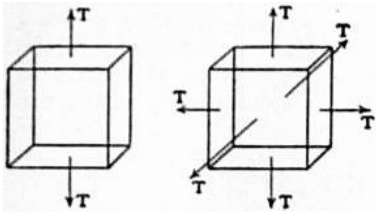

Fig. 1.

A state of stress in which there is purely tangential traction on a plane, and no normal traction on any perpendicular plane, is described as a state of “shearing stress.” The result (2) of § 6 shows that tangential tractions occur in pairs. If, at any point, there is tangential traction, in any direction, on a plane parallel to this direction, and if we draw through the point a plane at right angles to the direction of this traction, and therefore containing the normal to the first plane, then there is equal tangential traction on this second plane in the direction of the normal to the first plane. The result is illustrated in fig. 2, where a rectangular block is subjected on two opposite faces to opposing tangential tractions, and is held in equilibrium by equal tangential tractions applied to two other faces.

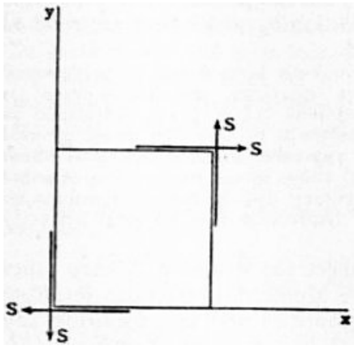

Fig. 2.

Through any point there always pass three planes, at right angles to each other, across which there is no tangential traction. These planes are called the “principal planes of stress,” and the (normal) tractions across them the “principal stresses.” Lines, usually curved, which have at every point the direction of a principal stress at the point, are called “lines of stress.”

8. It appears that the stress at any point of a body is completely specified by six quantities, which can be taken to be the X_x, Y_y, Z_z and Y_z, Z_x, X_y of § 6. The first three are tensions (pressures if they are negative) across three planes parallel to fixed rectangular directions, and the remaining three are tangential tractions across the same three planes. These six quantities are called the “components of stress.” It appears also that the components of stress are connected with each other, and with the body forces and accelerations, by the three partial differential equations of the type (3) of § 6. These equations are available for the purpose of determining the state of stress which exists in a body of definite form subjected to definite forces, but they are not sufficient for the purpose (see § 38 below). In order to effect the determination it is necessary to have information concerning the constitution of the body, and to introduce subsidiary relations founded upon this information.

9. The definite mathematical relations which have been found to connect the components of stress with each other, and with other quantities, result necessarily from the formation of a clear conception of the nature of stress. They do not admit of experimental verification, because the stress within a body does not admit of direct measurement. Results which are deduced by the aid of these relations can be compared with experimental results. If any discrepancy were observed it would not be interpreted as requiring a modification of the concept of stress, but as affecting some one or other of the subsidiary relations which must be introduced for the purpose of obtaining the theoretical result.

10. *Strain.*—For the specification of the changes of size and shape which are produced in a body by any forces, we begin by defining the “average extension” of any linear element or “filament” of the body. Let l_0 be the length of the filament before the forces are applied, l its length when the body is subjected to the forces. The average extension of the filament is measured by the fraction $(l - l_0)/l_0$. If this fraction is negative there is “contraction.” The “extension at a point” of a body in any assigned direction is the mathematical limit of this fraction when one end of the filament is at the point, the filament has the assigned direction, and its length is diminished indefinitely. It is clear that all the changes of size and shape of the body are known when the extension at every point in every direction is known.

The relations between the extensions in different directions around the same point are most simply expressed by introducing the extensions in the directions of the co-ordinate axes and the angles between filaments of the body which are initially parallel to these axes. Let e_{xx}, e_{yy}, e_{zz} denote the extensions parallel to the axes of x, y, z, and let e_{yz}, e_{zx}, e_{xy} denote the cosines of the angles between the pairs of filaments which are initially parallel to the axes of y and z, z and x, x and y. Also let e denote the extension in the direction of a line the direction cosines of which are l, m, n. Then, if the changes of size and shape are slight, we have the relation

$$e = e_{xx}l^2 + e_{yy}m^2 + e_{zz}n^2 + e_{yz}mn + e_{zx}nl + e_{xy}lm.$$

The body which undergoes the change of size or shape is said to be “strained,” and the “strain” is determined when the quantities e_{xx}, e_{yy}, e_{zz} and e_{yz}, e_{zx}, e_{xy} defined above are known at every point of it. These quantities are called “components of strain.” The three of the type e_{xx} are extensions, and the three of the type e_{yz} are called “shearing strains” (see § 12 below).

11. All the changes of relative position of particles of the body are known when the strain is known, and

conversely the strain can be determined when the changes of relative position are given. These changes can be expressed most simply by the introduction of a vector quantity to represent the displacement of any particle.

When the body is deformed by the action of any forces its particles pass from the positions which they occupied before the action of the forces into new positions. If x, y, z are the co-ordinates of the position of a particle in the first state, its co-ordinates in the second state may be denoted by x + u, y + v, z + w. The quantities, u, v, w are the “components of displacement.” When these quantities are small, the strain is connected with them by the equations

$$e_{xx} = \partial u / \partial x, \qquad e_{yy} = \partial v / \partial y, \qquad e_{zz} = \partial w / \partial z,$$
$$e_{yz} = \frac{\partial w}{\partial y} + \frac{\partial v}{\partial z}, \qquad e_{zx} = \frac{\partial u}{\partial z} + \frac{\partial w}{\partial x}, \qquad e_{xy} = \frac{\partial v}{\partial x} + \frac{\partial u}{\partial y}. \qquad (1)$$

12. These equations enable us to determine more exactly the nature of the “shearing strains” such as e_{xy}. Let u, for example, be of the form sy, where s is constant, and let v and w vanish. Then $e_{xy} = s$, and the remaining components of strain vanish. The nature of the strain (called “simple shear”) is simply appreciated by imagining the body to consist of a series of thin sheets, like the leaves of a book, which lie one over another and are all parallel to a plane (that of x, z); and the displacement is seen to consist in the shifting of each sheet relative to the sheet below in a direction (that of x) which is the same for all the sheets. The displacement of any sheet is proportional to its distance y from a particular sheet, which remains undisplaced. The shearing strain has the effect of distorting the shape of any portion of the body without altering its volume. This is shown in fig. 3, where a square ABCD is distorted by simple shear (each point moving parallel to the line marked xx) into a rhombus A′B′C′D′, as if by an extension of the diagonal BD and a contraction of the diagonal AC, which extension and contraction are adjusted so as to leave the area unaltered. In the general case, where u is not of the form sy and v and w do not vanish, the shearing strains such as e_{xy} result from the composition of pairs of simple shears of the type which has just been explained.

13. Besides enabling us to express the extension in any direction and the changes of relative direction of any filaments of the body, the components of strain also express the changes of size of volumes and areas. In particular, the “cubical dilatation,” that is to say, the increase of volume per unit of volume, is expressed by the quantity $e_{xx} + e_{yy} + e_{zz}$ or $\partial u / \partial x + \partial v / \partial y + \partial w / \partial z$. When this quantity is negative there is “compression.” 144

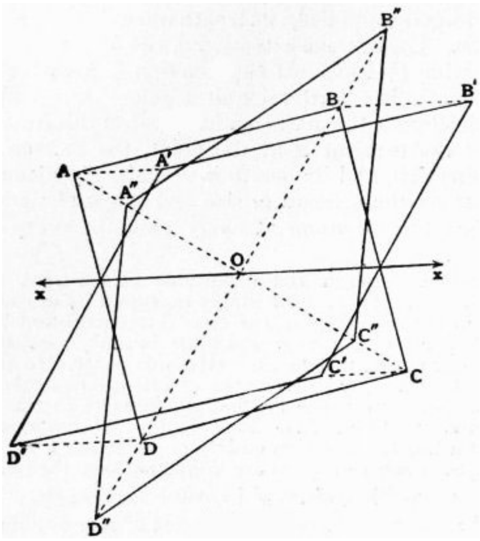

Fig. 3.

14. It is important to distinguish between two types of strain: the “rotational” type and the “irrotational” type. The distinction is illustrated in fig. 3, where the figure A″B″C″D″ is obtained from the figure ABCD by contraction parallel to AC and extension parallel to BD, and the figure A′B′C′D′ can be obtained from ABCD by the same contraction and extension followed by a rotation through the angle A″OA′. In strains of the irrotational type there are at any point three filaments at right angles to each other, which are such that the particles which lie in them before strain continue to lie in them after strain. A small spherical element of the body with its centre at the point becomes a small ellipsoid with its axes in the directions of these three filaments. In the case illustrated in the figure, the lines of the filaments in question, when the figure ABCD is strained into the figure A″B″C″D″, are OA, OB and a line through O at right angles to their plane. In strains of the rotational type, on the other hand, the single existing set of three filaments (issuing from a point) which cut each other at right angles both before and after strain do not retain their directions after strain, though one of them may do so in certain cases. In the figure, the lines of the filaments in question, when the figure ABCD is strained into A′B′C′D′, are OA, OB and a line at right angles to their plane before strain, and after strain they are OA′, OB′, and the same third line. A rotational strain can always be analysed into an irrotational strain (or “pure” strain) followed by a rotation.

Analytically, a strain is irrotational if the three quantities

$$\frac{\partial w}{\partial y} - \frac{\partial v}{\partial z}, \qquad \frac{\partial u}{\partial z} - \frac{\partial w}{\partial x}, \qquad \frac{\partial v}{\partial x} - \frac{\partial u}{\partial y}$$

vanish, rotational if any one of them is different from zero. The halves of these three quantities are the components of a vector quantity called the “rotation.”

15. Whether the strain is rotational or not, there is always one set of three linear elements issuing from any point which cut each other at right angles both before and after strain. If these directions are chosen as axes of x, y, z, the shearing strains e_{yz}, e_{zx}, e_{xy} vanish at this point. These directions are called the “principal axes of strain,” and the extensions in the directions of these axes the “principal extensions.”

16. It is very important to observe that the relations between components of strain and components of displacement imply relations between the components of strain themselves. If by any process of reasoning we arrive at the conclusion that the state of strain in a body is such and such a state, we have a test of the possibility or impossibility of our conclusion. The test is that, if the state of strain is a possible one, then there must be a displacement which can be associated with it in accordance with the equations (1) of § 11.

We may eliminate u, v, w from these equations. When this is done we find that the quantities e_{xx}, ... e_{yz} are connected by the two sets of equations

$$\begin{aligned} \frac{\partial^2 e_{yy}}{\partial z^2} + \frac{\partial^2 e_{zz}}{\partial y^2} &= \frac{\partial^2 e_{yz}}{\partial y \partial z} \\ \frac{\partial^2 e_{zz}}{\partial x^2} + \frac{\partial^2 e_{xx}}{\partial z^2} &= \frac{\partial^2 e_{zx}}{\partial z \partial x} \\ \frac{\partial^2 e_{xx}}{\partial y^2} + \frac{\partial^2 e_{yy}}{\partial x^2} &= \frac{\partial^2 e_{xy}}{\partial x \partial y} \end{aligned} \qquad (1)$$

and

$$\begin{aligned} 2\frac{\partial^2 e_{xx}}{\partial y \partial z} &= \frac{\partial}{\partial x}\left(-\frac{\partial e_{yz}}{\partial x} + \frac{\partial e_{zx}}{\partial y} + \frac{\partial e_{xy}}{\partial z} \right) \\ 2\frac{\partial^2 e_{yy}}{\partial z \partial x} &= \frac{\partial}{\partial y}\left(\frac{\partial e_{yz}}{\partial x} - \frac{\partial e_{zx}}{\partial y} + \frac{\partial e_{xy}}{\partial z} \right) \\ 2\frac{\partial^2 e_{zz}}{\partial x \partial y} &= \frac{\partial}{\partial z}\left(\frac{\partial e_{yz}}{\partial x} + \frac{\partial e_{zx}}{\partial y} - \frac{\partial e_{xy}}{\partial z} \right) \end{aligned} \qquad (2)$$

These equations are known as the *conditions of compatibility of strain-components*. The components of strain which specify any possible strain satisfy them. Quantities arrived at in any way, and intended to be components of strain, if they fail to satisfy these equations, are not the components of any possible strain; and the theory or speculation by which they are reached must be modified or abandoned.

When the components of strain have been found in accordance with these and other necessary equations, the displacement is to be found by solving the equations (1) of § 11, considered as differential equations to determine u, v, w. The most general possible solution will differ from any other solution by terms which contain arbitrary constants, and these terms represent a possible displacement. This “complementary displacement” involves no strain, and would be a possible displacement of an ideal perfectly rigid body.

17. The relations which connect the strains with each other and with the displacement are geometrical relations resulting from the definitions of the quantities and not requiring any experimental verification. They do not admit of such verification, because the strain within a body cannot be measured. The quantities (belonging to the same category) which can be measured are displacements of points on the surface of a body. For example, on the surface of a bar subjected to tension we may make two fine transverse scratches, and measure the distance between them before and after the bar is stretched. For such measurements very refined instruments are required. Instruments for this purpose are called barbarously “extensometers,” and many different kinds have been devised. From measurements of displacement by an extensometer we may deduce the average extension of a filament of the bar terminated by the two scratches. In general, when we attempt to measure a strain, we really measure some displacements, and deduce the values, not of the strain at a point, but of the average extensions of some particular linear filaments of a body containing the point; and these filaments are, from the nature of the case, nearly always superficial filaments.

18. In the case of transparent materials such as glass there is available a method of studying experimentally the state of strain within a body. This method is founded upon the result that a piece of glass when strained becomes doubly refracting, with its optical principal axes at any point in the directions of the principal axes of strain (§ 15) at the point. When the piece has two parallel plane faces, and two of the principal axes of strain at any point are parallel to these faces, polarized light transmitted through the piece in a direction normal to the faces can be used to determine the directions of the principal axes of the strain at any point. If the directions of these axes are known theoretically the comparison of the experimental and theoretical results yields a test of the theory.

19. *Relations between Stresses and Strains.*—The problem of the extension of a bar subjected to tension is the one which has been most studied experimentally, and as a result of this study it is found that for most materials, including all metals except cast metals, the measurable extension is proportional to the applied tension, provided that this tension is not too great. In interpreting this result it is assumed that the tension is uniform over the cross-section of the bar, and that the extension of longitudinal filaments is uniform throughout the bar; and then the result takes the form of a law of proportionality connecting stress and strain: The tension is proportional to the extension. Similar results are found for the same materials when other methods of experimenting are adopted, for example, when a bar is supported at the ends and bent by an attached load and the deflexion is measured, or when a bar is twisted by an axial couple and the relative angular displacement of two sections is measured. We have thus very numerous experimental verifications of the famous law first enunciated by Robert Hooke in 1678 in the words “*Ut Tensio sic vis*”; that is, “the Power of any spring is in the same proportion as the Tension (—stretching) thereof.” The most general statement of Hooke’s Law in modern language would be:—*Each of the six components of stress at any point of a body is a linear function of the six components of strain at the point.* It is evident from what has been said above as to the nature of the measurement of stresses and strains that this law in all its generality does not admit of complete experimental verification, and that the evidence for it consists largely in the agreement of the results which are deduced from it in a theoretical fashion with the results of experiments. Of such results one of a general character may be noted here. If the law is assumed to be true, and the equations of motion of the

body (§ 5) are transformed by means of it into differential equations for determining the components of displacement, these differential equations admit of solutions which represent periodic vibratory displacements (see § 85 below). The fact that solid bodies can be thrown into states of isochronous vibration has been emphasized by G.G. Stokes as a peremptory proof of the truth of Hooke's Law.

20. According to the statement of the generalized Hooke's Law the stress-components vanish when the strain-components vanish. The strain-components contemplated in experiments upon which the law is founded are measured from a zero of reckoning which corresponds to the state of the body subjected to experiment before the experiment is made, and the stress-components referred to in the statement of the law are those which are called into action by the forces applied to the body in the course of the experiment. No account is taken of the stress which must already exist in the body owing to the force of gravity and the forces by which the body is supported. When it is desired to take account of this stress it is usual to suppose that the strains which would be produced in the body if it could be freed from the action of gravity and from the pressures of supports are so small that the strains produced by the forces which are applied in the course of the experiment can be compounded with them by simple superposition. This supposition comes to the same thing as measuring the strain in the body, not from the state in which it was before the experiment, but from an ideal state (the "unstressed" state) in which it would be entirely free from internal stress, and allowing for the strain which would be produced by gravity and the supporting forces if these forces were applied to the body when free from stress. In most practical cases the initial strain to be allowed for is unimportant (see §§ 91-93 below).

21. Hooke's law of proportionality of stress and strain leads to the introduction of important physical constants: the *moduluses of elasticity* of a body. Let a bar of uniform section (of area ω) be stretched with tension T, which is distributed uniformly over the section, so that the stretching force is Twω, and let the bar be unsupported at the sides. The bar will undergo a longitudinal extension of magnitude T/E, where E is a constant quantity depending upon the material. This constant is called *Young's modulus* after Thomas Young, who introduced it into the science in 1807. The quantity E is of the same nature as a traction, that is to say, it is measured as a force estimated per unit of area. For steel it is about 2.04×10^{12} dynes per square centimetre, or about 13,000 tons per sq. in.

22. The longitudinal extension of the bar under tension is not the only strain in the bar. It is accompanied by a lateral contraction by which all the transverse filaments of the bar are shortened. The amount of this contraction is σT/E, where σ is a certain number called *Poisson's ratio*, because its importance was at first noted by S.D. Poisson in 1828. Poisson arrived at the existence of this contraction, and the corresponding number σ, from theoretical considerations, and his theory led him to assign to σ the value ¼. Many experiments have been made with the view of determining σ, with the result that it has been found to be different for different materials, although for very many it does not differ much from ¼. For steel the best value (Amagat's) is 0.268. Poisson's theory admits of being modified so as to agree with the results of experiment.

23. The behaviour of an elastic solid body, strained within the limits of its elasticity, is entirely determined by the constants E and σ if the body is *isotropic*, that is to say, if it has the same quality in all directions around any point. Nevertheless it is convenient to introduce other constants which are related to the action of particular sorts of forces. The most important of these are the "modulus of compression" (or "bulk modulus") and the "rigidity" (or "modulus of shear"). To define the *modulus of compression*, we suppose that a solid body of any form is subjected to uniform hydrostatic pressure of amount p. The state of stress within it will be one of uniform pressure, the same at all points, and the same in all directions round any point. There will be compression, the same at all points, and proportional to the pressure; and the amount of the compression can be expressed as p/k. The quantity k is the modulus of compression. In this case the linear contraction in any direction is p/3k; but in general the linear extension (or contraction) is not one-third of the cubical dilatation (or compression).

24. To define the *rigidity*, we suppose that a solid body is subjected to forces in such a way that there is shearing stress within it. For example, a cubical block may be subjected to opposing tractions on opposite faces acting in directions which are parallel to an edge of the cube and to both the faces. Let S be the amount of the traction, and let it be uniformly distributed over the faces. As we have seen (§ 7), equal tractions must act upon two other faces in suitable directions in order to maintain equilibrium (see fig. 2 of § 7). The two directions involved may be chosen as axes of x, y as in that figure. Then the state of stress will be one in which the stress-component denoted by X_y is equal to S, and the remaining stress-components vanish; and the strain produced in the body is shearing strain of the type denoted by e_{xy}. The amount of the shearing strain is S/μ, and the quantity μ is the "rigidity."

25. The modulus of compression and the rigidity are quantities of the same kind as Young's modulus. The modulus of compression of steel is about 1.43×10^{12} dynes per square centimetre, the rigidity is about 8.19×10^{11} dynes per square centimetre. It must be understood that the values for different specimens of nominally the same material may differ considerably.

The modulus of compression k and the rigidity μ of an isotropic material are connected with the Young's modulus E and Poisson's ratio σ of the material by the equations

$$k = E / 3(1 - 2\sigma), \quad \mu = E / 2(1 + \sigma).$$

26. Whatever the forces acting upon an isotropic solid body may be, provided that the body is strained within its limits of elasticity, the strain-components are expressed in terms of the stress-components by the equations

$$\begin{aligned} e_{xx} &= (X_x - \sigma Y_y - \sigma Z_z) / E, & e_{yz} &= Y_z / \mu, \\ e_{yy} &= (Y_y - \sigma Z_z - \sigma X_x) / E, & e_{zx} &= Z_x / \mu, \\ e_{zz} &= (Z_z - \sigma X_x - \sigma Y_y) / E, & e_{xy} &= X_y / \mu. \end{aligned} \quad (1)$$

If we introduce a quantity λ, of the same nature as E or μ, by the equation

$$\lambda = E\sigma / (1 + \sigma)(1 - 2\sigma), \qquad (2)$$

we may express the stress-components in terms of the strain-components by the equations

$$X_x = \lambda(e_{xx} + e_{yy} + e_{zz}) + 2\mu e_{xx}, \quad Y_z = \mu e_{yz},$$
$$Y_y = \lambda(e_{xx} + e_{yy} + e_{zz}) + 2\mu e_{yy}, \quad Z_x = \mu e_{zx},$$
$$Z_z = \lambda(e_{xx} + e_{yy} + e_{zz}) + 2\mu e_{zz}, \quad X_y = \mu e_{xy}; \qquad (3)$$

and then the behaviour of the body under the action of any forces depends upon the two constants λ and μ. 146
These two constants were introduced by G. Lamé in his treatise of 1852. The importance of the quantity μ had been previously emphasized by L.J. Vicat and G.G. Stokes.

27. The potential energy per unit of volume (often called the "resilience") stored up in the body by the strain is equal to

$$\tfrac{1}{2} (\lambda + 2\mu) (e_{xx} + e_{yy} + e_{zz})^2 + \tfrac{1}{2}\mu (e^2_{yz} + e^2_{zx} + e^2_{xy} - 4e_{yy}e_{zz} - 4e_{zz}e_{xx} - 4e_{xx}e_{yy}),$$

or the equivalent expression

$$\tfrac{1}{2} [(X^2_x + Y^2_y + Z^2_z) - 2\sigma (Y_yZ_z + Z_zX_x + X_xY_y) + 2 (1 + \sigma) (Y^2_z + Z^2_x + X^2_y)] / E.$$

The former of these expressions is called the "strain-energy-function."

28. The Young's modulus E of a material is often determined experimentally by the direct method of the extensometer (§ 17), but more frequently it is determined indirectly by means of a result obtained in the theory of the flexure of a bar (see §§ 47, 53 below). The rigidity μ is usually determined indirectly by means of results obtained in the theory of the torsion of a bar (see §§ 41, 42 below). The modulus of compression k may be determined directly by means of the piezometer, as was done by E.H. Amagat, or it may be determined indirectly by means of a result obtained in the theory of a tube under pressure, as was done by A. Mallock (see § 78 below). The value of Poisson's ratio σ is generally inferred from the relation connecting it with E and μ or with E and k, but it may also be determined indirectly by means of a result obtained in the theory of the flexure of a bar (§ 47 below), as was done by M.A. Cornu and A. Mallock, or directly by a modification of the extensometer method, as has been done recently by J. Morrow.

29. The *elasticity of a fluid* is always expressed by means of a single quantity of the same kind as the *modulus of compression* of a solid body. To any increment of pressure, which is not too great, there corresponds a proportional cubical compression, and the amount of this compression for an increment δp of pressure can be expressed as δp/k. The quantity that is usually tabulated is the reciprocal of k, and it is called the *coefficient of compressibility*. It is the amount of compression per unit increase of pressure. As a physical quantity it is of the same dimensions as the reciprocal of a pressure (or of a force per unit of area). The pressures concerned are usually measured in atmospheres (1 atmosphere = 1.014×10^6 dynes per sq. cm.). For water the coefficient of compressibility, or the compression per atmosphere, is about 4.5×10^{-5}. This gives for k the value 2.22×10^{10} dynes per sq. cm. The Young's modulus and the rigidity of a fluid are always zero.

30. The relations between stress and strain in a material which is not isotropic are much more complicated. In such a material the Young's modulus depends upon the direction of the tension, and its variations about a point are expressed by means of a surface of the fourth degree. The Poisson's ratio depends upon the direction of the contracted lateral filaments as well as upon that of the longitudinal extended ones. The rigidity depends upon both the directions involved in the specification of the shearing stress. In general there is no simple relation between the Young's moduluses and Poisson's ratios and rigidities for assigned directions and the modulus of compression. Many materials in common use, all fibrous woods for example, are actually *aeolotropic* (that is to say, are not isotropic), but the materials which are aeolotropic in the most regular fashion are natural crystals. The elastic behaviour of crystals has been studied exhaustively by many physicists, and in particular by W. Voigt. The strain-energy-function is a homogeneous quadratic function of the six strain-components, and this function may have as many as 21 independent coefficients, taking the place in the general case of the 2 coefficients λ, μ which occur when the material is isotropic—a result first obtained by George Green in 1837. The best experimental determinations of the coefficients have been made indirectly by Voigt by means of results obtained in the theories of the torsion and flexure of aeolotropic bars.

31. *Limits of Elasticity.*—A solid body which has been strained by considerable forces does not in general recover its original size and shape completely after the forces cease to act. The strain that is left is called *set*. If set occurs the elasticity is said to be "imperfect," and the greatest strain (or the greatest load) of any specified type, for which no set occurs, defines the "limit of perfect elasticity" corresponding to the specified type of strain, or of stress. All fluids and many solid bodies, such as glasses and crystals, as well as some metals (copper, lead, silver) appear to be perfectly elastic as regards change of volume within wide limits; but malleable metals and alloys can have their densities permanently increased by considerable pressures. The limits of perfect elasticity as regards change of shape, on the other hand, are very low, if they exist at all, for glasses and other hard, brittle solids; but a class of metals including copper, brass, steel, and platinum are very perfectly elastic as regards distortion, provided that the distortion is not too great. The question can be tested by observation of the torsional elasticity of thin fibres or wires. The limits of perfect elasticity are somewhat ill-defined, because an experiment cannot warrant us in asserting that there is no set, but only that, if there is any set, it is too small to be observed.

32. A different meaning may be, and often is, attached to the phrase "limits of elasticity" in consequence of the following experimental result:—Let a bar be held stretched under a moderate tension, and let the extension be measured; let the tension be slightly increased and the extension again measured; let this process be continued, the tension being increased by equal increments. It is found that when the tension is not too great the extension increases by equal increments (as nearly as experiment can decide), but that, as

the tension increases, a stage is reached in which the extension increases faster than it would do if it continued to be proportional to the tension. The beginning of this stage is tolerably well marked. Some time before this stage is reached the limit of perfect elasticity is passed; that is to say, if the load is removed it is found that there is some permanent set. The limiting tension beyond which the above law of proportionality fails is often called the "limit of *linear* elasticity." It is higher than the limit of perfect elasticity. For steel bars of various qualities J. Bauschinger found for this limit values varying from 10 to 17 tons per square inch. The result indicates that, when forces which produce any kind of strain are applied to a solid body and are gradually increased, the strain at any instant increases proportionally to the forces up to a stage beyond that at which, if the forces were removed, the body would completely recover its original size and shape, but that the increase of strain ceases to be proportional to the increase of load when the load surpasses a certain limit. There would thus be, for any type of strain, a *limit of linear elasticity*, which exceeds the limit of perfect elasticity.

33. A body which has been strained beyond the limit of linear elasticity is often said to have suffered an "over-strain." When the load is removed, the *set* which can be observed is not entirely permanent; but it gradually diminishes with lapse of time. This phenomenon is named "elastic after-working." If, on the other hand, the load is maintained constant, the strain is gradually increased. This effect indicates a gradual flowing of solid bodies under great stress; and a similar effect was observed in the experiments of H. Tresca on the punching and crushing of metals. It appears that all solid bodies under sufficiently great loads become "plastic," that is to say, they take a set which gradually increases with the lapse of time. No plasticity is observed when the limit of linear elasticity is not exceeded.

34. The values of the elastic limits are affected by overstrain. If the load is maintained for some time, and then removed, the limit of linear elasticity is found to be higher than before. If the load is not maintained, but is removed and then reapplied, the limit is found to be lower than before. During a period of rest a test piece recovers its elasticity after overstrain.

35. The effects of repeated loading have been studied by A. Wöhler, J. Bauschinger, O. Reynolds and others. It has been found that, after many repetitions of rather rapidly alternating stress, pieces are fractured by loads which they have many times withstood. It is not certain whether the fracture is in every case caused 147
by the gradual growth of minute flaws from the beginning of the series of tests, or whether the elastic quality of the material suffers deterioration apart from such flaws. It appears, however, to be an ascertained result that, so long as the limit of linear elasticity is not exceeded, repeated loads and rapidly alternating loads do not produce failure of the material.

36. The question of the conditions of safety, or of the conditions in which rupture is produced, is one upon which there has been much speculation, but no completely satisfactory result has been obtained. It has been variously held that rupture occurs when the numerically greatest principal stress exceeds a certain limit, or when this stress is tension and exceeds a certain limit, or when the greatest difference of two principal stresses (called the "stress-difference") exceeds a certain limit, or when the greatest extension or the greatest shearing strain or the greatest strain of any type exceeds a certain limit. Some of these hypotheses appear to have been disproved. It was held by G.F. Fitzgerald (*Nature*, Nov. 5, 1896) that rupture is not produced by pressure symmetrically applied all round a body, and this opinion has been confirmed by the recent experiments of A. Föppl. This result disposes of the greatest stress hypothesis and also of the greatest strain hypothesis. The fact that short pillars can be crushed by longitudinal pressure disposes of the greatest tension hypothesis, for there is no tension in the pillar. The greatest extension hypothesis failed to satisfy some tests imposed by H. Wehage, who experimented with blocks of wrought iron subjected to equal pressures in two directions at right angles to each other. The greatest stress-difference hypothesis and the greatest shearing strain hypothesis would lead to practically identical results, and these results have been held by J.J. Guest to accord well with his experiments on metal tubes subjected to various systems of combined stress; but these experiments and Guest's conclusion have been criticized adversely by O. Mohr, and the question cannot be regarded as settled. The fact seems to be that the conditions of rupture depend largely upon the nature of the test (tensional, torsional, flexural, or whatever it may be) that is applied to a specimen, and that no general formula holds for all kinds of tests. The best modern technical writings emphasize the importance of the limits of linear elasticity and of tests of dynamical resistance (§ 87 below) as well as of statical resistance.

37. The question of the conditions of rupture belongs rather to the science of the strength of materials than to the science of elasticity (§ 1); but it has been necessary to refer to it briefly here, because there is no method except the methods of the theory of elasticity for determining the state of stress or strain in a body subjected to forces. Whatever view may ultimately be adopted as to the relation between the conditions of safety of a structure and the state of stress or strain in it, the calculation of this state by means of the theory or by experimental means (as in § 18) cannot be dispensed with.

38. *Methods of determining the Stress in a Body subjected to given Forces.*—To determine the state of stress, or the state of strain, in an isotropic solid body strained within its limits of elasticity by given forces, we have to use (i.) the equations of equilibrium, (ii.) the conditions which hold at the bounding surface, (iii.) the relations between stress-components and strain-components, (iv.) the relations between strain-components and displacement. The equations of equilibrium are (with notation already used) three partial differential equations of the type

$$\frac{\partial X_x}{\partial x} + \frac{\partial X_y}{\partial y} + \frac{\partial Z_z}{\partial z} + \rho X = 0. \tag{1}$$

The conditions which hold at the bounding surface are three equations of the type

$$X_x \cos(x, \nu) + X_y \cos(y, \nu) + Z_x \cos(z, \nu) = \overline{X}_\nu, \tag{2}$$

where ν denotes the direction of the outward-drawn normal to the bounding surface, and $\overline{X}_\nu$ denotes the x-component of the applied surface traction. The relations between stress-components and strain-components are expressed by either of the sets of equations (1) or (3) of § 26. The relations between strain-components and displacement are the equations (1) of § 11, or the equivalent conditions of compatibility expressed in equations (1) and (2) of § 16.

39. We may proceed by either of two methods. In one method we eliminate the stress-components and the strain-components and retain only the components of displacement. This method leads (with notation already used) to three partial differential equations of the type

$$(\lambda + \mu) \frac{\partial}{\partial x} \left(\frac{\partial u}{\partial x} + \frac{\partial v}{\partial y} + \frac{\partial w}{\partial z} \right) + \mu \left(\frac{\partial^2 u}{\partial x^2} + \frac{\partial^2 u}{\partial y^2} + \frac{\partial^2 u}{\partial z^2} \right) + \rho X = 0, \quad (3)$$

and three boundary conditions of the type

$$\lambda \cos(x, \nu) \left(\frac{\partial u}{\partial x} + \frac{\partial v}{\partial y} + \frac{\partial w}{\partial z} \right) + \mu \left\{ 2 \cos(x, \nu) \frac{\partial u}{\partial x} + \cos(y, \nu) \left(\frac{\partial v}{\partial x} + \frac{\partial u}{\partial y} \right) + \cos(z, \nu) \left(\frac{\partial u}{\partial z} + \frac{\partial w}{\partial x} \right) \right\} = \overline{X}_\nu. \quad (4)$$

In the alternative method we eliminate the strain-components and the displacements. This method leads to a system of partial differential equations to be satisfied by the stress-components. In this system there are three equations of the type

$$\frac{\partial X_x}{\partial x} + \frac{\partial X_y}{\partial y} + \frac{\partial X_z}{\partial z} + \rho X = 0, \quad (1\ \textit{bis})$$

three of the type

$$\frac{\partial^2 X_x}{\partial x^2} + \frac{\partial^2 X_x}{\partial y^2} + \frac{\partial^2 X_x}{\partial z^2} + \frac{1}{1 + \sigma} \frac{\partial^2}{\partial x^2} (X_x + Y_y + Z_z) = - \frac{\sigma}{1 - \sigma} \rho \left(\frac{\partial X}{\partial x} + \frac{\partial Y}{\partial y} + \frac{\partial Z}{\partial z} \right) - 2\rho \frac{\partial X}{\partial x}, \quad (5)$$

and three of the type

$$\frac{\partial^2 Y_z}{\partial x^2} + \frac{\partial^2 Y_z}{\partial y^2} + \frac{\partial^2 Y_z}{\partial z^2} + \frac{1}{1 + \sigma} \frac{\partial^2}{\partial y \partial z} (X_x + Y_y + Z_z) = - \rho \left(\frac{\partial Z}{\partial y} + \frac{\partial Y}{\partial z} \right), \quad (6)$$

the equations of the two latter types being necessitated by the conditions of compatibility of strain-components. The solutions of these equations have to be adjusted so that the boundary conditions of the type (2) may be satisfied.

40. It is evident that whichever method is adopted the mathematical problem is in general very complicated. It is also evident that, if we attempt to proceed by help of some intuition as to the nature of the stress or strain, our intuition ought to satisfy the tests provided by the above systems of equations. Neglect of this precaution has led to many errors. Another source of frequent error lies in the neglect of the conditions in which the above systems of equations are correct. They are obtained by help of the supposition that the relative displacements of the parts of the strained body are small. The solutions of them must therefore satisfy the test of smallness of the relative displacements.

41. Torsion.—As a first example of the application of the theory we take the problem of the torsion of prisms. This problem, considered first by C.A. Coulomb in 1784, was finally solved by B. de Saint-Venant in 1855. The problem is this:—A cylindrical or prismatic bar is held twisted by terminal couples; it is required to determine the state of stress and strain in the interior. When the bar is a circular cylinder the problem is easy. Any section is displaced by rotation about the central-line through a small angle, which is proportional to the distance z of the section from a fixed plane at right angles to this line. This plane is a terminal section if one of the two terminal sections is not displaced. The angle through which the section z rotates is τz, where τ is a constant, called the amount of the twist; and this constant τ is equal to $G/\mu I$, where G is the twisting couple, and I is the moment of inertia of the cross-section about the central-line. This result is often called “Coulomb’s law.” The stress within the bar is shearing stress, consisting, as it must, of two sets of equal tangential tractions on two sets of planes which are at right angles to each other. These planes are the cross-sections and the axial planes of the bar. The tangential traction at any point of the cross-section is directed at right angles to the axial plane through the point, and the tangential traction on the axial plane is directed parallel to the length of the bar. The amount of either at a distance r from the axis is $\mu\tau r$ or Gr/I. The result that $G = \mu\tau I$ can be used to determine μ experimentally, for τ may be measured and G and I are known.

42. When the cross-section of the bar is not circular it is clear that this solution fails; for the existence of tangential traction, near the prismatic bounding surface, on any plane which does not cut this surface at right angles, implies the existence of traction applied to this surface. We may attempt to modify the theory by retaining the supposition that the stress consists of shearing stress, involving tangential traction distributed in some way over the cross-sections. Such traction is obviously a necessary constituent of any stress-system which could be produced by terminal couples around the axis. We should then know that there must be equal 148
tangential traction directed along the length of the bar, and exerted across some planes or other which are parallel to this direction. We should also know that, at the bounding surface, these planes must cut this surface at right angles. The corresponding strain would be shearing strain which could involve (i.) a sliding of elements of one cross-section relative to another, (ii.) a relative sliding of elements of the above mentioned planes in the direction of the length of the bar. We could conclude that there may be a longitudinal displacement of the elements of the cross-sections. We should then attempt to satisfy the conditions of the problem by supposing that this is the character of the strain, and that the corresponding displacement consists of (i.) a rotation of the cross-sections in their planes such as we found in the case of the circle, (ii.) a distortion of the cross-sections into curved surfaces by a displacement (w) which is directed normally to their planes and varies in some manner from point to point of these planes. We could show that all the conditions of the problem are satisfied by this assumption, provided that the longitudinal displacement (w), considered

as a function of the position of a point (x, y) in the cross-section, satisfies the equation

$$\frac{\partial^2 w}{\partial x^2} + \frac{\partial^2 w}{\partial y^2} = 0, \qquad (1)$$

and the boundary condition

$$\left(\frac{\partial w}{\partial x} - \tau y\right)\cos(x, \nu) + \left(\frac{\partial w}{\partial y} + \tau x\right)\cos(y, \nu) = 0, \qquad (2)$$

where τ denotes the amount of the twist, and ν the direction of the normal to the boundary. The solution is known for a great many forms of section. (In the particular case of a circular section w vanishes.) The tangential traction at any point of the cross-section is directed along the tangent to that curve of the family ψ = const. which passes through the point, ψ being the function determined by the equations

$$\frac{\partial w}{\partial x} = \tau\left(\frac{\partial \psi}{\partial y} + y\right), \quad \frac{\partial w}{\partial y} = -\tau\left(\frac{\partial \psi}{\partial x} + x\right).$$

The amount of the twist τ produced by terminal couples of magnitude G is G/C, where C is a constant, called the "torsional rigidity" of the prism, and expressed by the formula

$$C = \mu \iint \left\{\left(\frac{\partial \psi}{\partial x}\right)^2 + \left(\frac{\partial \psi}{\partial y}\right)^2\right\} dxdy,$$

the integration being taken over the cross-section. When the coefficient of μ in the expression for C is known for any section, μ can be determined by experiment with a bar of that form of section.

43. The distortion of the cross-sections into curved surfaces is shown graphically by drawing the contour lines (w = const.). In general the section is divided into a number of compartments, and the portions that lie within two adjacent compartments are respectively concave and convex. This result is illustrated in the accompanying figures (fig. 4 for the ellipse, given by $x^2/b^2 + y^2/c^2 = 1$; fig. 5 for the equilateral triangle, given by $(x + \frac{1}{3}a)(x^2 - 3y^2 - \frac{4}{3}ax + \frac{4}{9}a^2) = 0$; fig. 6 for the square).

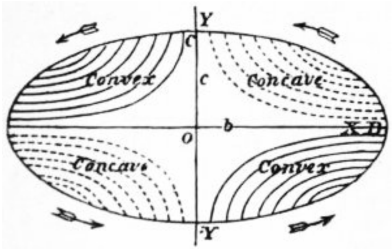

Fig. 4.

44. The distribution of the shearing stress over the cross-section is determined by the function ψ, already introduced. If we draw the curves ψ = const., corresponding to any form of section, for equidifferent values of the constant, the tangential traction at any point on the cross-section is directed along the tangent to that curve of the family which passes through the point, and the magnitude of it is inversely proportional to the distance between consecutive curves of the family. Fig. 7 illustrates the result in the case of the *equilateral* triangle. The boundary is, of course, one of the lines. The "lines of shearing stress" which can thus be drawn are in every case identical with the lines of flow of frictionless liquid filling a cylindrical vessel of the same cross-section as the bar, when the liquid circulates in the plane of the section with uniform spin. They are also the same as the contour lines of a flexible and slightly extensible membrane, of which the edge has the same form as the bounding curve of the cross-section of the bar, when the membrane is fixed at the edge and slightly deformed by uniform pressure.

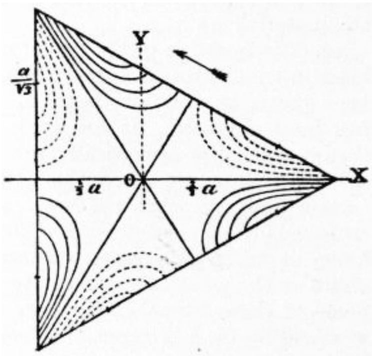

Fig. 5.

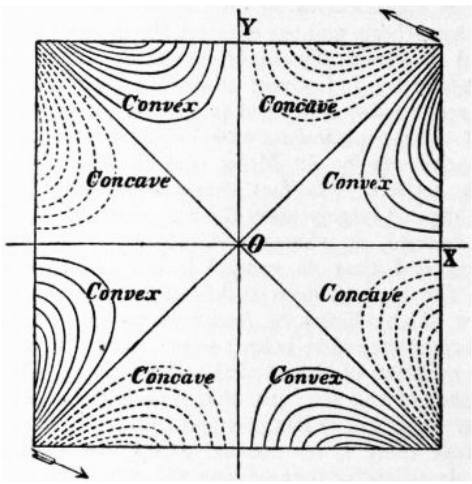

Fig. 6.

45. Saint-Venant's theory shows that the true torsional rigidity is in general less than that which would be obtained by extending Coulomb's law ($G = \mu\tau I$) to sections which are not circular. For an elliptic cylinder of sectional area ω and moment of inertia I about its central-line the torsional rigidity is $\mu\omega^4 / 4\pi^2 I$, and this formula is not far from being correct for a very large number of sections. For a bar of square section of side a centimetres, the torsional rigidity in C.G.S.

units is (0.1406) μa^4 approximately, μ being expressed in dynes per square centimetre. How great the defect of the true value from that given by extending Coulomb’s law may be in the case of sections with projecting corners is shown by the diagrams (fig. 8 especially no. 4). In these diagrams the upper of the two numbers under each figure indicates the fraction which the true torsional rigidity corresponding to the section is of that value which would be obtained by extending Coulomb’s law; and the lower of the two numbers indicates the ratio which the torsional rigidity for a bar of the corresponding section bears to that of a bar of circular section of the same material and of equal sectional area.

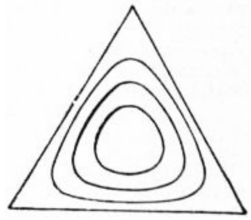

Fig. 7.

These results have an important practical application, inasmuch as they show that strengthening ribs and projections, such as are introduced in engineering to give stiffness to beams, have the reverse of a good effect when torsional stiffness is an object, although they are of great value in increasing the resistance to bending. The theory shows further that the resistance to torsion is very seriously diminished when there is in the surface any dent approaching to a re-entrant angle. At such a place the shearing strain tends to become infinite, and some permanent set is produced by torsion. In the case of a 149 section of any form, the strain and stress are greatest at points on the contour, and these points are in many cases the points of the contour which are nearest to the centroid of the section. The theory has also been applied to show that a longitudinal flaw near the axis of a shaft transmitting a torsional couple has little influence on the strength of the shaft, but that in the neighbourhood of a similar flaw which is much nearer to the surface than to the axis the shearing strain may be nearly doubled, and thus the possibility of such flaws is a source of weakness against which special provision ought to be made.

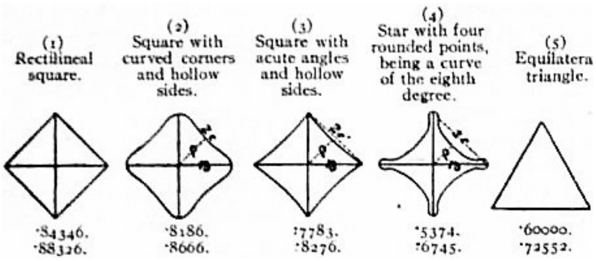

Fig. 8.—Diagrams showing Torsional Rigidities.

46. *Bending of Beams.*—As a second example of the application of the general theory we take the problem of the flexure of a beam. In this case also we begin by forming a simple intuition as to the nature of the strain and the stress. On the side of the beam towards the centre of curvature the longitudinal filaments must be contracted, and on the other side they must be extended. If we assume that the cross-sections remain plane, and that the central-line is unaltered in length, we see (at once from fig. 9) that the extensions (or contractions) are given by the formula y/R, where y denotes the distance of a longitudinal filament from the plane drawn through the unstrained central-line at right-angles to the plane of bending, and R is the radius of curvature of the curve into which this line is bent (shown by the dotted line in the figure). Corresponding to this strain there must be traction acting across the cross-sections. If we assume that there is no other stress, then the magnitude of the traction in question is Ey/R, where E is Young’s modulus, and it is tension on the side where the filaments are extended and pressure on the side where they are contracted. If the plane of bending contains a set of principal axes of the cross-sections at their centroids, these tractions for the whole cross-section are equivalent to a couple of moment EI/R, where I now denotes the moment of inertia of the cross-section about an axis through its centroid at right angles to the plane of bending, and the plane of the couple is the plane of bending. Thus a beam of any form of section can be held bent in a “principal plane” by terminal couples of moment M, that is to say by a “bending moment” M; the central-line will take a curvature M/EI, so that it becomes an arc of a circle of radius EI/M; and the stress at any point will be tension of amount My/I, where y denotes distance (reckoned positive towards the side remote from the centre of curvature) from that plane which initially contains the central-line and is at right angles to the plane of the couple. This plane is called the “neutral plane.” The restriction that the beam is bent in a principal plane means that the plane of bending contains one set of principal axes of the cross-sections at their centroids; in the case of a beam of rectangular section the plane would bisect two opposite edges at right angles. In order that the theory may hold good the radius of curvature must be very large.

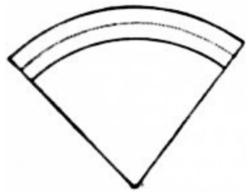

Fig. 9.

47. In this problem of the bending of a beam by terminal couples the stress is tension, determined as above, and the corresponding strain consists therefore of longitudinal extension of amount My/EI or y/R (contraction if y is negative), accompanied by lateral contraction of amount $\sigma My/EI$ or $\sigma y/R$ (extension if y is negative), σ being Poisson’s ratio for the material. Our intuition of the nature of the strain was imperfect, inasmuch as it took no account of these lateral strains. The necessity for introducing them was pointed out by Saint-Venant. The effect of them is a change of shape of the cross-sections in their own planes. This is shown in an exaggerated way in fig. 10, where the rectangle ABCD represents the cross-section of the unstrained beam, or a rectangular portion of this cross-section, and the curvilinear figure A′B′C′D′ represents in an exaggerated fashion the cross-section (or the corresponding portion of the cross-section) of the same beam, when bent so that the centre of curvature of the central-line (which is at right angles to the plane

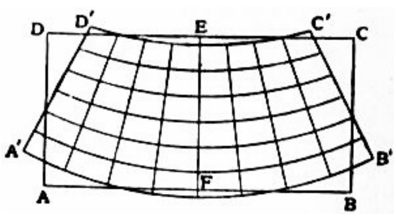

Fig. 10.

of the figure) is on the line EF produced beyond F. The lines A′B′ and C′D′ are approximately circles of radii R/σ, when the central-line is a circle of radius R, and their centres are on the line FE produced beyond E. Thus the neutral plane, and each of the faces that is parallel to it, becomes strained into an *anticlastic surface*, whose principal curvatures are in the ratio $\sigma : 1$. The general appearance of the bent beam is shown in an exaggerated fashion in fig. 11, where the traces of the surface into which the neutral plane is bent are dotted. The result that the ratio of the principal curvatures of the anticlastic surfaces, into which the top and bottom planes of the beam (of rectangular section) are bent, is Poisson's ratio σ, has been used for the experimental determination of σ. The result that the radius of curvature of the bent central-line is EI/M is used in the experimental determination of E. The quantity EI is often called the "flexural rigidity" of the beam. There are two principal flexural rigidities corresponding to bending in the two principal planes (cf. § 62 below).

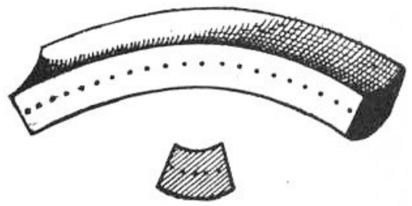

Fig. 11.

48. That this theory requires modification, when the load does not consist simply of terminal couples, can be seen most easily by considering the problem of a beam loaded at one end with a weight W, and supported in a horizontal position at its other end. The forces that are exerted at any section p, to balance the weight W, must reduce statically to a vertical force W and a couple, and these forces arise from the action of the part Ap on the part Bp (see fig. 12), *i.e.* from the stresses across the section at p. The couple is equal to the moment of the applied load W about an axis drawn through the centroid of the section p at right angles to the plane of bending. This moment is called the "bending moment" at the section, it is the product of the load W and the distance of the section from the loaded end, so that it varies uniformly along the length of the beam. The stress that suffices in the simpler problem gives rise to no vertical force, and it is clear that in addition to longitudinal tensions and pressures there must be tangential tractions on the cross-sections. The resultant of these tangential tractions must be a force equal to W, and directed vertically; but the direction of 150
the traction at a point of the cross-section need not in general be vertical. The existence of tangential traction on the cross-sections implies the existence of equal tangential traction, directed parallel to the central-line, on some planes or other which are parallel to this line, the two sets of tractions forming a shearing stress. We conclude that such shearing stress is a necessary constituent of the stress-system in the beam bent by terminal transverse load. We can develop a theory of this stress-system from the assumptions (i.) that the tension at any point of the cross-section is related to the bending moment at the section by the same law as in the case of uniform bending by terminal couples; (ii.) that, in addition to this tension, there is at any point shearing stress, involving tangential tractions acting in appropriate directions upon the elements of the cross-sections. When these assumptions are made it appears that there is one and only one distribution of shearing stress by which the conditions of the problem can be satisfied. The determination of the amount and direction of this shearing stress, and of the corresponding strains and displacements, was effected by Saint-Venant and R.F.A. Clebsch for a number of forms of section by means of an analysis of the same kind as that employed in the solution of the torsion problem.

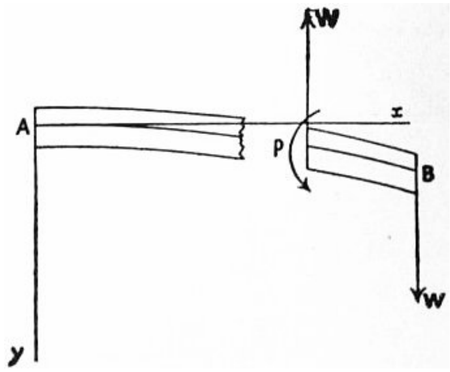

Fig. 12.

49. Let l be the length of the beam, x the distance of the section p from the fixed end A, y the distance of any point below the horizontal plane through the centroid of the section at A, then the bending moment at p is $W(l - x)$, and the longitudinal tension P or X_x at any point on the cross-section is $-W(l - x)y/I$, and this is related to the bending moment exactly as in the simpler problem.

50. The expressions for the shearing stresses depend on the shape of the cross-section. Taking the beam to be of isotropic material and the cross-section to be an ellipse of semiaxes a and b (fig. 13), the a axis being vertical in the unstrained state, and drawing the axis z at right angles to the plane of flexure, we find that the vertical shearing stress U or X_y at any point (y, z) on any cross-section is

$$\frac{2W[(a^2 - y^2)\{2a^2(1 + \sigma) + b^2\} - z^2a^2(1 - 2\sigma)]}{\pi a^3 b(1 + \sigma)(3a^2 + b^2)}.$$

The resultant of these stresses is W, but the amount at the centroid, which is the maximum amount, exceeds the average amount, $W/\pi ab$, in the ratio

$$\{4a^2(1 + \sigma) + 2b^2\} / (3a^2 + b^2)(1 + \sigma).$$

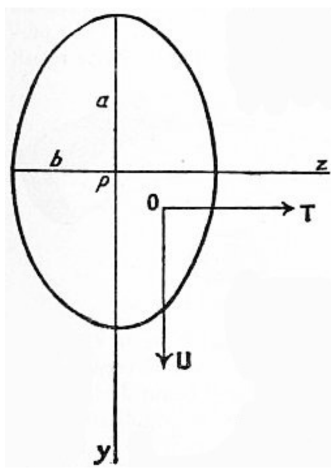

Fig. 13.

If $\sigma = \frac{1}{4}$, this ratio is $\frac{7}{5}$ for a circle, nearly $\frac{4}{3}$ for a flat elliptic bar with the longest diameter vertical, nearly $\frac{8}{5}$ for a flat elliptic bar with the longest diameter horizontal.

In the same problem the horizontal shearing stress T or Z_x at any point on any cross-section is of amount

$$-\frac{4Wyz\{a^2(1 + \sigma) + b^2\sigma\}}{\pi a^3 b(1 + \sigma)(3a^2 + b^2)}.$$

The resultant of these stresses vanishes; but, taking as before $\sigma = \frac{1}{4}$, and putting for the three cases above $a = b$, $a = 10b$, $b = 10a$, we find that the ratio of the maximum of this stress to the average vertical shearing stress has the values $\frac{3}{5}$, nearly $\frac{1}{15}$, and nearly 4. Thus the stress T is of considerable importance when the beam is a plank.

As another example we may consider a circular tube of external radius r_0 and internal radius r_1. Writing P, U, T for X_x, X_y, Z_x, we find

$$P = -\frac{4W}{\pi(r_0^4 - r_1^4)}(l - x)y,$$

$$U = \frac{W}{2(1+\sigma)\pi(r_0^4 - r_1^4)}\left[(3+2\sigma)\left\{r_0^2 + r_1^2 - y^2 - \frac{r_0^2 r_1^2}{(y^2+z^2)^2}(y^2 - z^2)\right\} - (1-2\sigma)z^2\right]$$

$$T = -\frac{W}{(1+\sigma)\pi(r_0^4 - r_1^4)}\left\{1 + 2\sigma + (3+2\sigma)\frac{r_0^2 r_1^2}{(y^2+z^2)^2}\right\}yz;$$

and for a tube of radius r and small thickness t the value of P and the maximum values of U and T reduce approximately to

$$P = -W(l - x)y / \pi r^3 t$$

$$U_{max.} = W / \pi rt, \quad T_{max.} = W / 2\pi rt.$$

The greatest value of U is in this case approximately twice its average value, but it is possible that these results for the bending of very thin tubes may be seriously at fault if the tube is not plugged, and if the load is not applied in the manner contemplated in the theory (cf. § 55). In such cases the extensions and contractions of the longitudinal filaments may be practically confined to a small part of the material near the ends of the tube, while the rest of the tube is deformed without stretching.

51. The tangential tractions U, T on the cross-sections are necessarily accompanied by tangential tractions on the longitudinal sections, and on each such section the tangential traction is parallel to the central line; on a vertical section z = const. its amount at any point is T, and on a horizontal section y = const. its amount at any point is U.

The internal stress at any point is completely determined by the components P, U, T, but these are not principal stresses (§ 7). Clebsch has given an elegant geometrical construction for determining the principal stresses at any point when the values of P, U, T are known.

From the point O (fig. 14) draw lines OP, OU, OT, to represent the stresses P, U, T at O, on the cross-section through O, in magnitude, direction and sense, and compound U and T into a resultant represented by OE; the plane EOP is a principal plane of stress at O, and the principal stress at right angles to this plane vanishes. Take M the middle point of OP, and with centre M and radius ME describe a circle cutting the line OP in A and B; then OA and OB represent the magnitudes of the two remaining principal stresses. On AB describe a rectangle ABDC so that DC passes through E; then OC is the direction of the principal stress represented in magnitude by OA, and OD is the direction of the principal stress represented in magnitude by OB.

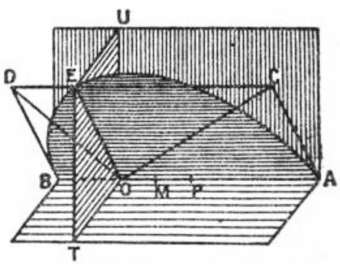

Fig. 14.

52. As regards the strain in the beam, the longitudinal and lateral extensions and contractions depend on the bending moment in the same way as in the simpler problem; but, the bending moment being variable, the anticlastic curvature produced is also variable. In addition to these extensions and contractions there are shearing strains corresponding to the shearing stresses T, U. The shearing strain corresponding to T consists of a relative sliding parallel to the central-line of different longitudinal linear elements combined with a relative sliding in a transverse horizontal direction of elements of different cross-sections; the latter of these is concerned in the production of those displacements by which the variable anticlastic curvature is brought about; to see the effect of the former we may most suitably consider, for the case of an elliptic cross-section, the distortion of the shape of a rectangular portion of a plane of the material which in the natural state was horizontal; all the boundaries of such a portion become parabolas of small curvature, which is variable along the length of the beam, and the particular effect under consideration is the change of the transverse horizontal linear elements from straight lines such as HK to parabolas such as H′K′ (fig. 15); the lines HL and KM are parallel to the central-line, and the figure is drawn for a plane above the neutral plane. When the cross-section is not an ellipse the character of the strain is the same, but the curves are only approximately parabolic.

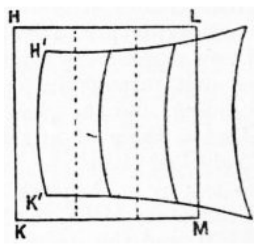

Fig. 15.

The shearing strain corresponding to U is a distortion which has the effect that the straight vertical filaments become curved lines which cut the longitudinal filaments obliquely, and thus the cross-sections do not remain plane, but become curved surfaces, and the tangent plane to any one of these surfaces at the centroid cuts the central line obliquely (fig. 16). The angle between these tangent planes and the central-line is the same at all points of the line; and, if it is denoted by $\frac{1}{2}\pi + s_0$, the value of s_0 is expressible as

$$\frac{\text{shearing stress at centroid}}{\text{rigidity of material}},$$

and it thus depends on the shape of the cross-section; for the elliptic section of § 50 its value is

$$\frac{4W}{E\pi ab}\ \frac{2a^2(1+\sigma) + b^2}{3a^2 + b^2};$$

for a circle (with σ = ¼) this becomes $7W / 2E\pi a^2$. The vertical filament through the centroid of any cross-section becomes a cubical parabola, as shown in fig. 16, and the contour lines of the curved surface into which any cross-section is distorted are shown in fig. 17 for a circular section.

53. The deflection of the beam is determined from the equation

curvature of central line = bending moment ÷ flexural rigidity, and the special conditions at the supported end; there is no alteration of this statement on account of the shears. As regards the special condition at an end which is *encastrée*, or built in, Saint-Venant proposed to assume that the central tangent plane of the cross-section at the end is vertical; with this assumption the tangent to the central line at the end is inclined downwards and makes an angle s_0 with the horizontal (see fig. 18); it is, however, improbable that this condition is exactly realized in practice. In the application of the theory to the experimental determination of Young’s modulus, the small angle which the central-line at the support makes with the horizontal is an unknown quantity, to be eliminated by observation of the deflection at two or more points.

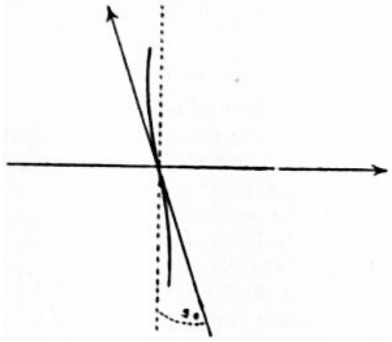

Fig. 16.

54. We may suppose the displacement in a bent beam to be produced by the following operations: (1) the central-line is deflected into its curved form, (2) the cross-sections are rotated about axes through their centroids at right angles to the plane of flexure so as to make angles equal to $\frac{1}{2}\pi + s_0$ with the central-line, (3) each cross-section is distorted in its own plane in such a way that the appropriate variable anticlastic curvature is produced, (4) the cross-sections are further distorted into curved surfaces. The contour lines of fig. 17 show the disturbance from the central tangent plane, not from the original vertical plane.

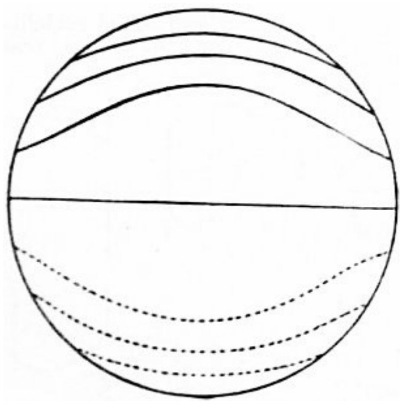

Fig. 17.

55. *Practical Application of Saint-Venant’s Theory.*—The theory above described is exact provided the forces applied to the loaded end, which have W for resultant, are distributed over the terminal section in a particular way, not likely to be realized in practice; and the application to practical problems depends on a principle due to Saint-Venant, to the effect that, except for comparatively small portions of the beam near to the loaded and fixed ends, the resultant only is effective, and its mode of distribution does not seriously affect the internal strain and stress. In fact, the actual stress is that due to forces with the required resultant distributed in the manner contemplated in the theory, superposed upon that due to a certain distribution of forces on each terminal section which, if applied to a rigid body, would keep it in equilibrium; according to Saint-Venant’s principle, the stresses and strains due to such distributions of force are unimportant except near the ends. For this principle to be exactly applicable it is necessary that the length of the beam should be very great compared with any linear dimension of its cross-section; for the practical application it is sufficient that the length should be about ten times the greatest diameter.

56. In recent years the problem of the bending of a beam by loads distributed along its length has been much advanced. It is now practically solved for the case of a load distributed uniformly, or according to any rational algebraic law, and it is also solved for the case where the thickness is small compared with the length and depth, as in a plate girder, and the load is distributed in any way. These solutions are rather complicated and difficult to interpret. The case which has been worked out most fully is that of a transverse load distributed uniformly along the length of the beam. In this case two noteworthy results have been obtained. The first of these is that the central-line in general suffers extension. This result had been found experimentally many years before. In the case of the plate girder loaded uniformly along the top, this extension is just half as great as the extension of the central-line of the same girder when free at the ends, supported along the base, and carrying the same load along the top. The second noteworthy result is that the curvature of the strained central-line is not proportional to the bending moment. Over and above the curvature which would be found from the ordinary relation—

curvature of central-line = bending moment ÷ flexural rigidity,

there is an additional curvature which is the same at all the cross-sections. In ordinary cases, provided the length is large compared with any linear dimension of the cross-section, this additional curvature is small compared with that calculated from the ordinary formula, but it may become important in cases like that of suspension bridges, where a load carried along the middle of the roadway is supported by tensions in rods attached at the sides.

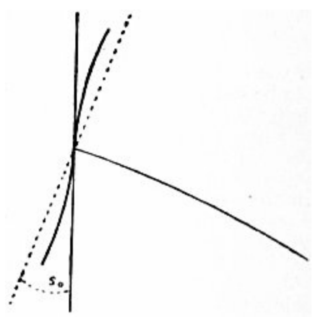

Fig. 18.

57. When the ordinary relation between the curvature and the bending moment is applied to the calculation of the deflection of *continuous beams* it must not be forgotten that a correction of the kind just mentioned may possibly be requisite. In the usual method of treating the problem such corrections are not considered, and the ordinary relation is made the basis of the theory. In order to apply this relation to the calculation of the deflection, it is necessary to know the bending moment at every point; and, since the pressures of the supports are not among the data of the problem, we require a method of determining the bending moments at the supports either by calculation or in some other way. The calculation of the bending moment can be replaced by a method of graphical construction, due to Mohr, and depending on the two following theorems:—

(i.) The curve of the central-line of each span of a beam, when the bending moment M is given,[1] is identical with the catenary or funicular curve passing through the ends of the span under a (fictitious) load per unit length of the span equal to M/EI, the horizontal tension in the funicular being unity.

(ii.) The directions of the tangents to this funicular curve at the ends of the span are the same for all statically equivalent systems of (fictitious) load.

When M is known, the magnitude of the resultant shearing stress at any section is dM/dx, where x is measured along the beam.

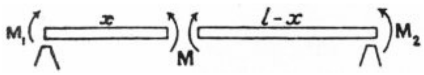

FIG. 20.

58. Let l be the length of a span of a loaded beam (fig. 19), M_1 and M_2 the bending moments at the ends, M the bending moment at a section distant x from the end (M_1), M′ the bending moment at the same section when the same span with the same load is simply supported; then M is given by the formula

$$M = M' + M_1 \frac{l - x}{l} + M_2 \frac{x}{l},$$

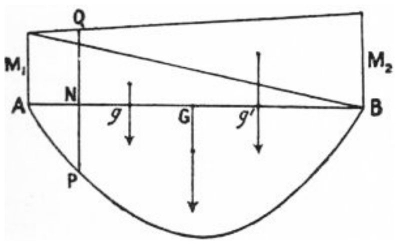

FIG. 19.

and thus a fictitious load statically equivalent to M/EI can be easily found when M′ has been found. If we draw a curve (fig. 20) to pass through the ends of the span, so that its ordinate represents the value of M′/EI, the corresponding fictitious loads are statically equivalent to a single load, of amount represented by the area of the curve, placed at the point of the span vertically above the centre of gravity of this area. If PN is the ordinate of this curve, and if at the ends of the span we erect ordinates in the proper sense to represent M_1/EI and M_2/EI, the bending moment at any point is represented by the length PQ.[2] For a uniformly distributed load the curve of M′ is a parabola M′ = ½wx (l − x), where w is the load per unit of length; and the statically equivalent fictitious load is $\frac{1}{12}wl^3$ / EI placed at the middle point G of the span; also the loads statically equivalent to the fictitious loads M_1 (l − x) / lEI and M_2x / lEI are ½M_1l / EI and ½M_2l / EI placed at the points g, g′ of trisection of the span. The funicular polygon for the fictitious loads can thus be drawn, and the direction of the central-line at the supports is determined when the bending moments at the supports are known.

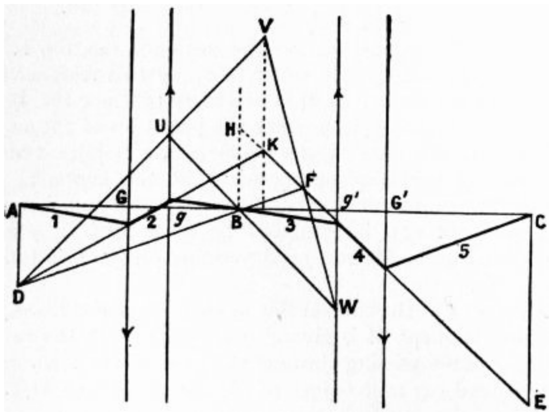

FIG. 21.

59. When there is more than one span the funiculars in question may be drawn for each of the spans, and, if the bending moments at the ends of the extreme spans are known, the intermediate ones can be determined. This determination depends on two considerations: (1) the fictitious loads corresponding to the bending moment at any support are proportional to the lengths of the spans which abut on that support; (2) the sides of two funiculars that end at any support coincide in direction. Fig. 21 illustrates the method for the case of a uniform beam on three supports A, B, C, the ends A and C being freely supported. There will be an unknown bending moment M_0 at B, and the system[3] of fictitious loads is $\frac{1}{12}wAB^3$/EI at G the middle point of AB, $\frac{1}{12}wBC^3$ / EI at G′ the middle point of BC, −½M_0AB / EI at g and −½M_0BC / EI at g′, where g and g′ are the points of trisection nearer to B of the spans AB, BC. The centre of gravity of the two latter is a fixed point independent of M_0, and the line VK of the figure is the vertical through this point. We draw AD and CE to represent the loads at G and G′ in magnitude; then D and E are fixed points. We construct any triangle UVW whose sides UV, UW pass through D, B, and whose vertices lie on the verticals gU, VK, g′W; the point F where VW meets DB is a fixed point, and the lines EF, DK are the two sides (2, 4) of the required funiculars which do not pass through A, B or C. The remaining sides (1, 3, 5) can then be drawn, and the side 3 necessarily passes through B; for the triangle UVW and the triangle whose sides are 2, 3, 4 are in perspective.

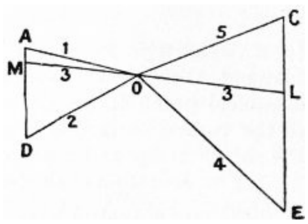

FIG. 22.

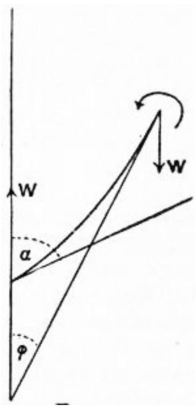

FIG. 23.

The bending moment M_0 is represented in the figure by the vertical line BH where H is on the continuation of the side 4, the scale being given by

BH ½M_0BC

$\overline{\text{CE}} = \overline{\tfrac{1}{12}\text{wBC}^3}$;

this appears from the diagrams of forces, fig. 22, in which the oblique lines are marked to correspond to the sides of the funiculars to which they are parallel.

In the application of the method to more complicated cases there are two systems of fixed points corresponding to F, by means of which the sides of the funiculars are drawn.

60. *Finite Bending of Thin Rod.*—The equation

$$\text{curvature} = \text{bending moment} \div \text{flexural rigidity}$$

may also be applied to the problem of the flexure in a principal plane of a very thin rod or wire, for which the curvature need not be small. When the forces that produce the flexure are applied at the ends only, the curve into which the central-line is bent is one of a definite family of curves, to which the name *elastica* has been given, and there is a division of the family into two species according as the external forces are applied directly to the ends or are applied to rigid arms attached to the ends; the curves of the former species are characterized by the presence of inflections at all the points at which they cut the line of action of the applied forces.

We select this case for consideration. The problem of determining the form of the curve (cf. fig. 23) is mathematically identical with the problem of determining the motion of a simple circular pendulum oscillating through a finite angle, as is seen by comparing the differential equation of the curve

$$\text{EI}\,\frac{d^2\varphi}{ds^2} + \text{W} \sin \varphi = 0$$

with the equation of motion of the pendulum

$$l\,\frac{d^2\varphi}{dt^2} + g \sin \varphi = 0.$$

The length L of the curve between two inflections corresponds to the time of oscillation of the pendulum from rest to rest, and we thus have

$$\text{L}\,\sqrt{(\text{W} / \text{EI})} = 2\text{K},$$

where K is the real quarter period of elliptic functions of modulus sin $\tfrac{1}{2}\alpha$, and α is the angle at which the curve cuts the line of action of the applied forces. Unless the length of the rod exceeds $\pi\sqrt{(\text{EI} / \text{W})}$ it will not bend under the force, but when the length is great enough there may be more than two points of inflection and more than one bay of the curve; for n bays (n + 1 inflections) the length must exceed $n\pi\sqrt{(\text{EI} / \text{W})}$. Some of the forms of the curve are shown in fig. 24.

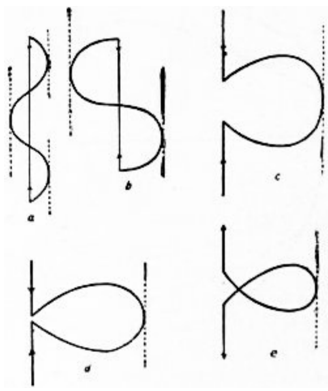

Fig. 24.

For the form d, in which two bays make a figure of eight, we have

$$\text{L}\sqrt{(\text{W} / \text{EI})} = 4.6, \qquad \alpha = 130°$$

approximately. It is noteworthy that whenever the length and force admit of a sinuous form, such as a or b, with more than two inflections, there is also possible a crossed form, like e, with two inflections only; the latter form is stable and the former unstable.

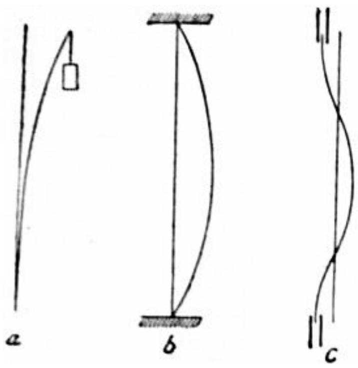

Fig. 25.

61. The particular case of the above for which α is very small is a curve of sines of small amplitude, and the result in this case has been applied to the problem of the buckling of struts under thrust. When the strut, of length L′, is maintained upright at its lower 153
end, and loaded at its upper end, it is simply contracted, unless $\text{L}'^2\text{W} > \tfrac{1}{4}\pi^2\text{EI}$, for the lower end corresponds to a point at which the tangent is vertical on an elastica for which the line of inflections is also vertical, and thus the length must be half of one bay (fig. 25, a). For greater lengths or loads the strut tends to bend or buckle under the load. For a very slight excess of $\text{L}'^2\text{W}$ above $\tfrac{1}{4}\pi^2\text{EI}$, the theory on which the above discussion is founded, is not quite adequate, as it assumes the central-line of the strut to be free from extension or contraction, and it is probable that bending without extension does not take place when the length or the force exceeds the critical value but slightly. It should be noted also that the formula has no application to short struts, as the theory from which it is derived is founded on the assumption that the length is great compared with the diameter (cf. § 56).

The condition of buckling, corresponding to the above, for a long strut, of length L′, when both ends are free to turn is $\text{L}'^2\text{W} > \pi^2\text{EI}$; for the central-line forms a complete bay (fig. 25, b); if both ends are maintained in the same vertical line, the condition is $\text{L}'^2\text{W} > 4\pi^2\text{EI}$, the central-line forming a complete bay and two half bays (fig. 25, *c*).

62. In our consideration of flexure it has so far been supposed that the bending takes place in a principal plane. We may remove this restriction by resolving the forces that tend to produce bending into systems of forces acting in the two principal planes. To each plane there corresponds a particular flexural rigidity, and the systems of forces in the two planes give rise to independent systems of stress, strain and displacement, which must be superposed in order to obtain the actual state. Applying this process to the problem of §§ 48-54, and supposing that one principal axis of a cross-section at its centroid makes an angle θ with the vertical, then for any shape of section the neutral surface or locus of unextended fibres cuts the section in a line DD′,

which is conjugate to the vertical diameter CP with respect to any ellipse of inertia of the section. The central-line is bent into a plane curve which is not in a vertical plane, but is in a plane through the line CY which is perpendicular to DD′ (fig. 26).

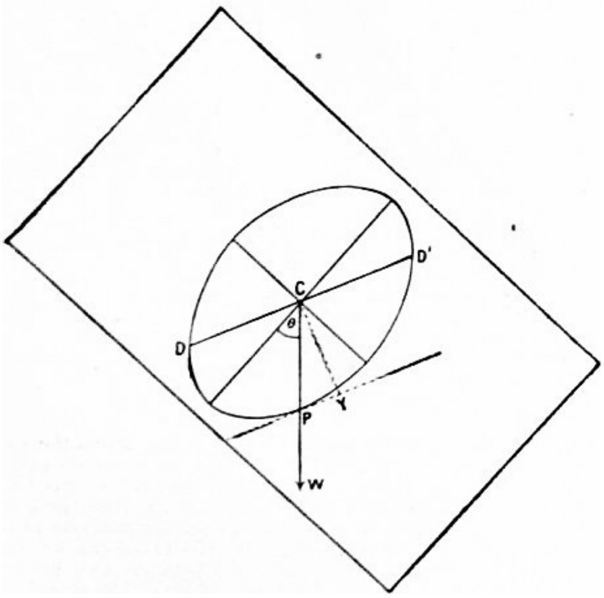

FIG. 26.

63. *Bending and Twisting of Thin Rods.*—When a very thin rod or wire is bent and twisted by applied forces, the forces on any part of it limited by a normal section are balanced by the tractions across the section, and these tractions are statically equivalent to certain forces and couples at the centroid of the section; we shall call them the *stress-resultants* and the *stress-couples*. The stress-couples consist of two flexural couples in the two principal planes, and the torsional couple about the tangent to the central-line. The torsional couple is the product of the torsional rigidity and the twist produced; the torsional rigidity is exactly the same as for a straight rod of the same material and section twisted without bending, as in Saint-Venant's torsion problem (§ 42). The twist τ is connected with the deformation of the wire in this way: if we suppose a very small ring which fits the cross-section of the wire to be provided with a pointer in the direction of one principal axis of the section at its centroid, and to move along the wire with velocity v, the pointer will rotate about the central-line with angular velocity τv. The amount of the flexural couple for either principal plane at any section is the product of the flexural rigidity for that plane, and the resolved part in that plane of the curvature of the central line at the centroid of the section; the resolved part of the curvature along the normal to any plane is obtained by treating the curvature as a vector directed along the normal to the osculating plane and projecting this vector. The flexural couples reduce to a single couple in the osculating plane proportional to the curvature when the two flexural rigidities are equal, and in this case only.

The stress-resultants across any section are tangential forces in the two principal planes, and a tension or thrust along the central-line; when the stress-couples and the applied forces are known these stress-resultants are determinate. The existence, in particular, of the resultant tension or thrust parallel to the central-line does not imply sensible extension or contraction of the central filament, and the tension per unit area of the cross-section to which it would be equivalent is small compared with the tensions and pressures in longitudinal filaments not passing through the centroid of the section; the moments of the latter tensions and pressures constitute the flexural couples.

64. We consider, in particular, the case of a naturally straight spring or rod of circular section, radius c, and of homogeneous isotropic material. The torsional rigidity is $\frac{1}{4}E\pi c^4 / (1 + \sigma)$; and the flexural rigidity, which is the same for all planes through the central-line, is $\frac{1}{4}E\pi c^4$; we shall denote these by C and A respectively. The rod may be held bent by suitable forces into a curve of double curvature with an amount of twist τ, and then the torsional couple is Cτ, and the flexural couple in the osculating plane is A/ρ, where ρ is the radius of circular curvature. Among the curves in which the rod can be held by forces and couples applied at its ends only, one is a circular helix; and then the applied forces and couples are equivalent to a wrench about the axis of the helix.

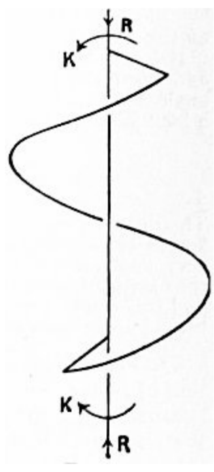

FIG. 27.

Let α be the angle and r the radius of the helix, so that ρ is $r \sec^2\alpha$; and let R and K be the force and couple of the wrench (fig. 27).

Then the couple formed by R and an equal and opposite force at any section and the couple K are equivalent to the torsional and flexural couples at the section, and this gives the equations for R and K

$$R = A \frac{\sin\alpha \cos^3\alpha}{r^2} - \frac{\cos\alpha}{r},$$

$$K = A \frac{\cos^3\alpha}{r} + C\tau \sin\alpha.$$

The thrust across any section is R sin α parallel to the tangent to the helix, and the shearing stress-resultant is R cos α at right angles to the osculating plane.

When the twist is such that, if the rod were simply unbent, it would also be untwisted, τ is $(\sin\alpha\cos\alpha)/r$, and then, restoring the values of A and C, we have

$$R = \frac{E\pi c^4}{4r^2}\,\frac{\sigma}{1+\sigma}\sin\alpha\cos^2\alpha,$$
$$K = \frac{E\pi c^4}{4r}\,\frac{1+\sigma\cos^2\alpha}{1+\sigma}\cos\alpha.$$

65. The theory of spiral springs affords an application of these results. The stress-couples called into play when a naturally helical spring (α, r) is held in the form of a helix (α', r'), are equal to the differences between those called into play when a straight rod of the same material and section is held in the first form, and those called into play when it is held in the second form.

Thus the torsional couple is

$$C\left(\frac{\sin\alpha'\cos\alpha'}{r'} - \frac{\sin\alpha\cos\alpha}{r}\right).$$

and the flexural couple is

$$A\left(\frac{\cos^2\alpha'}{r'} - \frac{\cos^2\alpha}{r}\right).$$

The wrench (R, K) along the axis by which the spring can be held in the form (α', r') is given by the equations

$$R = A\,\frac{\sin\alpha'}{r'}\left(\frac{\cos^2\alpha'}{r'} - \frac{\cos^2\alpha}{r}\right) - C\,\frac{\cos\alpha'}{r'}\left(\frac{\sin\alpha'\cos\alpha'}{r'} - \frac{\sin\alpha\cos\alpha}{r}\right),$$
$$K = A\cos\alpha'\left(\frac{\cos^2\alpha'}{r'} - \frac{\cos^2\alpha}{r}\right) + \sin\alpha'\left(\frac{\sin\alpha'\cos\alpha'}{r'} - \frac{\sin\alpha\cos\alpha}{r}\right).$$

When the spring is slightly extended by an axial force F, $= -R$, and there is no couple, so that K vanishes, and α', r' differ very little from α, r, it follows from these equations that the axial elongation, δx, is connected with the axial length x and the force F by the equation

$$F = \frac{E\pi c^4}{4r^2}\,\frac{\sin\alpha}{1+\sigma\cos^2\alpha}\,\frac{\delta x}{x},$$

and that the loaded end is rotated about the axis of the helix through a small angle

$$\frac{4\sigma F x r\cos\alpha}{E\pi c^4},$$

the sense of the rotation being such that the spring becomes more tightly coiled.

66. A horizontal pointer attached to a vertical spiral spring would be made to rotate by loading the spring, and the angle through which it turns might be used to measure the load, at any rate, when the load is not too great; but a much more sensitive contrivance is the twisted strip devised by W.E. Ayrton and J. Perry. A very thin, narrow rectangular strip of metal is given a permanent twist about its longitudinal middle line, and a pointer is attached to it at right angles to this line. When the strip is subjected to longitudinal tension the pointer rotates through a considerable angle. G.H. Bryan (*Phil. Mag.*, December 1890) has succeeded in constructing a theory of the action of the strip, according to which it is regarded as a strip of *plating* in the form of a right helicoid, which, after extension of the middle line, becomes a portion of a slightly different helicoid; on account of the thinness of the strip, the change of curvature of the surface is considerable, even when the extension is small, and the pointer turns with the generators of the helicoid.

If b stands for the breadth and t for the thickness of the strip, and τ for the permanent twist, the approximate formula for the angle θ through which the strip is untwisted on the application of a load W was found to be

$$\theta = \frac{Wb\tau\,(1+\sigma)}{2Et^3\left(1 + \frac{(1+\sigma)}{30}\,\frac{b^4\tau^2}{t^2}\right)}.$$

The quantity $b\tau$ which occurs in the formula is the total twist in a length of the strip equal to its breadth, and this will generally be very small; if it is small of the same order as t/b, or a higher order, the formula becomes $\tfrac{1}{2}Wb\tau(1+\sigma)/Et^3$, with sufficient approximation, and this result appears to be in agreement with observations of the behaviour of such strips.

67. *Thin Plate under Pressure.*—The theory of the deformation of plates, whether plane or curved, is very intricate, partly because of the complexity of the kinematical relations involved. We shall here indicate the nature of the effects produced in a thin plane plate, of isotropic material, which is slightly bent by pressure. This theory should have an application to the stress produced in a ship’s plates. In the problem of the cylinder under internal pressure (§ 77 below) the most important stress is the circumferential tension, counteracting the tendency of the circular filaments to expand under the pressure; but in the problem of a plane plate some of the filaments parallel to the plane of the plate are extended and others are contracted, so that the tensions and pressures along them give rise to resultant couples but not always to resultant forces. Whatever forces are applied to bend the plate, these couples are always expressible, at least approximately in terms of the principal curvatures produced in the surface which, before strain, was the middle plane of the plate. The simplest case is that of a rectangular plate, bent by a distribution of couples applied to its edges, so that the middle surface becomes a cylinder of large radius R; the requisite couple per unit of length of the straight edges is of amount C/R, where C is a certain constant; and the requisite couple per unit of length of the circular edges is of amount $C\sigma/R$, the latter being required to resist the tendency to anticlastic curvature (cf. § 47). If normal sections of the plate are supposed drawn through the generators and circular sections of the cylinder, the action of the neighbouring portions on any portion so bounded involves flexural couples of the above amounts. When the plate is bent in any manner, the curvature produced at each section of the middle surface may be regarded as arising from the superposition of two cylindrical curvatures; and the

flexural couples across normal sections through the lines of curvature, estimated per unit of length of those lines, are $C(1/R_1 + \sigma/R_2)$ and $C(1/R_2 + \sigma/R_1)$, where R_1 and R_2 are the principal radii of curvature. The value of C for a plate of small thickness 2h is $\frac{2}{3}Eh^3 / (1 - \sigma^2)$. Exactly as in the problem of the beam (§§ 48, 56), the action between neighbouring portions of the plate generally involves shearing stresses across normal sections as well as flexural couples; and the resultants of these stresses are determined by the conditions that, with the flexural couples, they balance the forces applied to bend the plate.

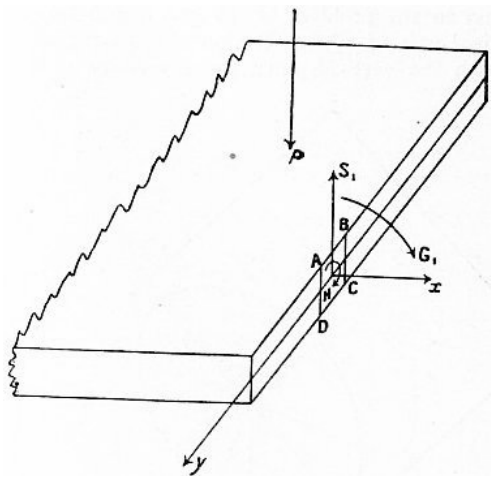

FIG. 28.

68. To express this theory analytically, let the middle plane of the plate in the unstrained position be taken as the plane of (x, y), and let normal sections at right angles to the axes of x and y be drawn through any point. After strain let w be the displacement of this point in the direction perpendicular to the plane, marked p in fig. 28. If the axes of x and y were parallel to the lines of curvature at the point, the flexural couple acting across the section normal to x (or y) would have the axis of y (or x) for its axis; but when the lines of curvature are inclined to the axes of co-ordinates, the flexural couple across a section normal to either axis has a component about that axis as well as a component about the perpendicular axis. Consider an element ABCD of the section at right angles to the axis of x, contained between two lines near together and perpendicular to the middle plane. The action of the portion of the plate to the right upon the portion to the left, across the element, gives rise to a couple about the middle line (y) of amount, estimated per unit of length of that line, equal to $C[\partial^2 w/\partial x^2 + \sigma(\partial^2 w/\partial y^2)]$, $= G_1$, say, and to a couple, similarly estimated, about the normal (x) of amount $-C(1 - \sigma)(\partial^2 w/\partial x \partial y)$, H, say. The corresponding couples on an element of a section at 155
right angles to the axis of y, estimated per unit of length of the axis of x, are of amounts $-C[\partial^2 w/\partial y^2 + \sigma(\partial^2 w/\partial x^2)]$, $= G_2$ say, and −H. The resultant S_1 of the shearing stresses on the element ABCD, estimated as before, is given by the equation $S_1 = \partial G_1/\partial x - \partial H/\partial y$ (cf. § 57), and the corresponding resultant S_2 for an element perpendicular to the axis of y is given by the equation $S_2 = -\partial H/\partial x - \partial G_2/\partial y$. If the plate is bent by a pressure p per unit of area, the equation of equilibrium is $\partial S_1/\partial x + \partial S_2/\partial y = p$, or, in terms of w,

$$\frac{\partial^4 w}{\partial x^4} + \frac{\partial^4 w}{\partial y^4} + 2\frac{\partial^4 w}{\partial x^2 \partial y^2} = \frac{p}{C}.$$

This equation, together with the special conditions at the rim, suffices for the determination of w, and then all the quantities here introduced are determined. Further, the most important of the stress-components are those which act across elements of normal sections: the tension in direction x, at a distance z from the middle plane measured in the direction of p, is of amount $-3Cz/2h^3 [\partial^2 w/\partial x^2 + \sigma(\partial^2 w/\partial y^2)]$, and there is a corresponding tension in direction y; the shearing stress consisting of traction parallel to y on planes x = const., and traction parallel to x on planes y = const., is of amount $[3C(1 - \sigma)z/2h^3] \cdot (\partial^2 w/\partial x \partial y)$; these tensions and shearing stresses are equivalent to two principal tensions, in the directions of the lines of curvature of the surface into which the middle plane is bent, and they give rise to the flexural couples.

69. In the special example of a circular plate, of radius a, supported at the rim, and held bent by a uniform pressure p, the value of w at a point distant r from the axis is

$$\frac{1}{64}\frac{p}{C}(a^2 - r^2)\left(\frac{5 + \sigma}{1 + \sigma}a^2 - r^2\right),$$

and the most important of the stress components is the radial tension, of which the amount at any point is $\frac{3}{32}(3 + \sigma)\,pz\,(a^2 - r)/h^3$; the maximum radial tension is about $\frac{1}{3}(a/h)^2 p$, and, when the thickness is small compared with the diameter, this is a large multiple of p.

70. *General Theorems.*—Passing now from these questions of flexure and torsion, we consider some results that can be deduced from the general equations of equilibrium of an elastic solid body.

The form of the general expression for the potential energy (§ 27) stored up in the strained body leads, by a general property of quadratic functions, to a reciprocal theorem relating to the effects produced in the body by two different systems of forces, viz.: The whole work done by the forces of the first system, acting over the displacements produced by the forces of the second system, is equal to the whole work done by the forces of the second system, acting over the displacements produced by the forces of the first system. By a suitable choice of the second system of forces, the average values of the component stresses and strains produced by given forces, considered as constituting the first system, can be obtained, even when the distribution of the stress and strain cannot be determined.

Taking for example the problem presented by an isotropic body of any form[4] pressed between two parallel planes distant l apart (fig. 29), and denoting the resultant pressure by p, we find that the diminution of volume $-\delta v$ is given by the equation

$$-\delta v = lp / 3k,$$

where k is the modulus of compression, equal to $\frac{1}{3}E / (1 - 2\sigma)$. Again, take the problem of the changes produced in a heavy body by different ways of supporting it; when the body is suspended from one or more points in a horizontal plane its volume is increased by

$$\delta v = Wh / 3k,$$

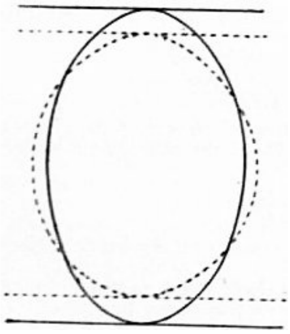

FIG. 29.

where W is the weight of the body, and h the depth of its centre of gravity below the plane; when the body is supported by upward vertical pressures at one or more points in a horizontal plane the volume is diminished by

$$-\delta v = Wh' / 3k,$$

where h′ is the height of the centre of gravity above the plane; if the body is a cylinder, of length l and section A, standing with its base on a smooth horizontal plane, its length is shortened by an amount

$$-\delta l = Wl / 2EA;$$

if the same cylinder lies on the plane with its generators horizontal, its length is increased by an amount

$$\delta l = \sigma Wh' / EA.$$

71. In recent years important results have been found by considering the effects produced in an elastic solid by forces applied at isolated points.

Taking the case of a single force F applied at a point in the interior, we may show that the stress at a distance r from the point consists of

(1) a radial pressure of amount

$$\frac{2-\sigma}{1-\sigma} \frac{F}{4\pi} \frac{\cos\theta}{r^2},$$

(2) tension in all directions at right angles to the radius of amount

$$\frac{1-2\sigma}{2(1-\sigma)} \frac{F}{4\pi} \frac{\cos\theta}{r^2},$$

(3) shearing stress consisting of traction acting along the radius *dr* on the surface of the cone θ = const. and traction acting along the meridian dθ on the surface of the sphere r = const. of amount

$$\frac{1-2\sigma}{2(1-\sigma)} \frac{F}{4\pi} \frac{\sin\theta}{r^2},$$

where θ is the angle between the radius vector r and the line of action of F. The line marked T in fig. 30 shows the direction of the tangential traction on the spherical surface.

Thus the lines of stress are in and perpendicular to the meridian plane, and the direction of one of those in the meridian plane is inclined to the radius vector r at an angle

$$\tfrac{1}{2}\tan^{-1}\left(\frac{2-4\sigma}{5-4\sigma}\tan\theta\right).$$

The corresponding displacement at any point is compounded of a radial displacement of amount

$$\frac{1+\sigma}{2(1-\sigma)} \frac{F}{4\pi E} \frac{\cos\theta}{r}$$

and a displacement parallel to the line of action of F of amount

$$\frac{(3-4\sigma)(1+\sigma)}{2(1-\sigma)} \frac{F}{4\pi E} \frac{1}{r}.$$

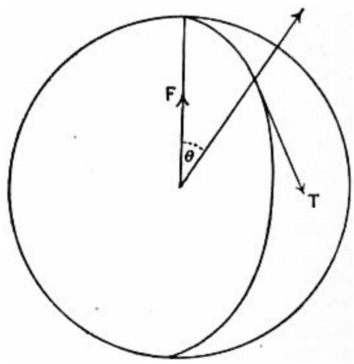

FIG. 30.

The effects of forces applied at different points and in different directions can be obtained by summation, and the effect of continuously distributed forces can be obtained by integration.

72. The stress system considered in § 71 is equivalent, on the plane through the origin at right angles to the line of action of F, to a resultant pressure of magnitude ½F at the origin and a $[1 - 2\sigma/2(1 - \sigma)] \cdot F/4\pi r^2$, and, by the application of this system of tractions to a solid bounded by a plane, the displacement just described would be produced. There is also another stress system for a solid so bounded which is equivalent, on the same plane, to a resultant pressure at the origin, and a radial traction proportional to $1/r^2$, but these are in the ratio $2\pi : r^{-2}$, instead of being in the ratio $4\pi(1 - \sigma) : (1 - 2\sigma)r^{-2}$.

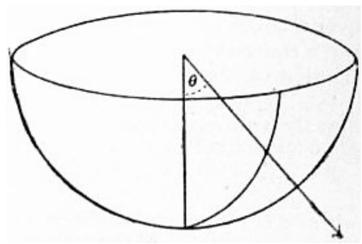

FIG. 31.

The second stress system (see fig. 31) consists of:

(1) radial pressure $F'r^{-2}$,

(2) tension in the meridian plane across the radius vector of amount

$$F'r^{-2} \cos\theta / (1 + \cos\theta),$$

(3) tension across the meridian plane of amount

$$F'r^{-2} / (1 + \cos\theta),$$

(4) shearing stress as in § 71 of amount

$$F'r^{-2} \sin\theta / (1 + \cos\theta),$$

and the stress across the plane boundary consists of a resultant pressure of magnitude $2\pi F'$ and a radial traction of amount $F'r^{-2}$. If then we superpose the component stresses of the last section multiplied by $4(1 - \sigma)W/F$, and the component stresses here written down multiplied by $-(1 - 2\sigma)W/2\pi F'$, the stress on the plane boundary will reduce to a single pressure W at the origin. We shall thus obtain the stress system at any point due to such a force applied at one point of the boundary. 156

In the stress system thus arrived at the traction across any plane parallel to the boundary is directed away from the place where W is supported, and its amount is $3W \cos^2\theta / 2\pi r^2$. The corresponding displacement consists of

(1) a horizontal displacement radially outwards from the vertical through the origin of amount

$$\frac{W (1 + \sigma) \sin\theta}{2\pi E r} \left(\cos\theta - \frac{1 - 2\sigma}{1 + \cos\theta} \right),$$

(2) a vertical displacement downwards of amount

$$\frac{W (1 + \sigma)}{2\pi E r} \{2 (1 - \sigma) + \cos^2\theta \}.$$

The effects produced by a system of loads on a solid bounded by a plane can be deduced.

The results for a solid body bounded by an infinite plane may be interpreted as giving the local effects of forces applied to a small part of the surface of a body. The results show that pressure is transmitted into a body from the boundary in such a way that the traction at a point on a section parallel to the boundary is the same at all points of any sphere which touches the boundary at the point of pressure, and that its amount at any point is inversely proportional to the square of the radius of this sphere, while its direction is that of a line drawn from the point of pressure to the point at which the traction is estimated. The transmission of force through a solid body indicated by this result was strikingly demonstrated in an attempt that was made to measure the lunar deflexion of gravity; it was found that the weight of the observer on the floor of the laboratory produced a disturbance of the instrument sufficient to disguise completely the effect which the instrument had been designed to measure (see G.H. Darwin, *The Tides and Kindred Phenomena in the Solar System*, London, 1898).

73. There is a corresponding theory of two-dimensional systems, that is to say, systems in which either the displacement is parallel to a fixed plane, or there is no traction across any plane of a system of parallel planes. This theory shows that, when pressure is applied at a point of the edge of a plate in any direction in the plane of the plate, the stress developed in the plate consists exclusively of radial pressure across any circle having the point of pressure as centre, and the magnitude of this pressure is the same at all points of any circle which touches the edge at the point of pressure, and its amount at any point is inversely proportional to the radius of this circle. This result leads to a number of interesting solutions of problems relating to plane systems; among these may be mentioned the problem of a circular plate strained by any forces applied at its edge.

74. The results stated in § 72 have been applied to give an account of the nature of the actions concerned in the impact of two solid bodies. The dissipation of energy involved in the impact is neglected, and the resultant pressure between the bodies at any instant during the impact is equal to the rate of destruction of momentum of either along the normal to the plane of contact drawn towards the interior of the other. It has been shown that in general the bodies come into contact over a small area bounded by an ellipse, and remain in contact for a time which varies inversely as the fifth root of the initial relative velocity.

For equal spheres of the same material, with $\sigma = \frac{1}{4}$, impinging directly with relative velocity v, the patches that come into contact are circles of radius

$$\left(\frac{45\pi}{256} \right)^{\frac{1}{5}} \left(\frac{v}{V} \right)^{\frac{2}{5}} r,$$

where r is the radius of either, and V the velocity of longitudinal waves in a thin bar of the material. The duration of the impact is approximately

$$(2.9432) \left(\frac{2025\pi^2}{512} \right)^{\frac{1}{5}} \frac{r}{v^{1/5} V^{4/5}}.$$

For two steel spheres of the size of the earth impinging with a velocity of 1 cm. per second the duration of the impact would be about twenty-seven hours. The fact that the duration of impact is, for moderate velocities, a considerable multiple of the time taken by a wave of compression to travel through either of two impinging bodies has been ascertained experimentally, and constitutes the reason for the adequacy of the statical theory here described.

75. *Spheres and Cylinders.*—Simple results can be found for spherical and cylindrical bodies strained by radial forces.

For a sphere of radius a, and of homogeneous isotropic material of density ρ, strained by the mutual gravitation of its parts, the stress at a distance r from the centre consists of

(1) uniform hydrostatic pressure of amount $\frac{1}{10} g\rho a (3 - \sigma) / (1 - \sigma)$,

(2) radial tension of amount $\frac{1}{10} g\rho (r^2/a) (3 - \sigma) / (1 - \sigma)$,

(3) uniform tension at right angles to the radius vector of amount

$$\tfrac{1}{10}\, g\rho\, (r^2/a)\, (1 + 3\sigma) / (1 - \sigma),$$

where g is the value of gravity at the surface. The corresponding strains consist of

(1) uniform contraction of all lines of the body of amount

$$\tfrac{1}{30}\, k^{-1} g\rho a\, (3 - \sigma) / (1 - \sigma),$$

(2) radial extension of amount $\tfrac{1}{10}\, k^{-1} g\rho\, (r^2/a)\, (1 + \sigma) / (1 - \sigma)$,

(3) extension in any direction at right angles to the radius vector of amount

$$\tfrac{1}{30}\, k^{-1} g\rho\, (r^2/a)\, (1 + \sigma) / (1 - \sigma),$$

where k is the modulus of compression. The volume is diminished by the fraction $g\rho a/5k$ of itself. The parts of the radii vectors within the sphere $r = a\, \{(3 - \sigma) / (3 + 3\sigma)\}^{1/2}$ are contracted, and the parts without this sphere are extended. The application of the above results to the state of the interior of the earth involves a neglect of the caution emphasized in § 40, viz. that the strain determined by the solution must be small if the solution is to be accepted. In a body of the size and mass of the earth, and having a resistance to compression and a rigidity equal to those of steel, the radial contraction at the centre, as given by the above solution, would be nearly ⅓, and the radial extension at the surface nearly ⅙, and these fractions can by no means be regarded as "small."

76. In a spherical shell of homogeneous isotropic material, of internal radius r_1 and external radius r_0, subjected to pressure p_0 on the outer surface, and p_1 on the inner surface, the stress at any point distant r from the centre consists of

(1) uniform tension in all directions of amount

$$\frac{p_1 r_1^3 - p_0 r_0^3}{r_0^3 - r_1^3},$$

(2) radial pressure of amount

$$\frac{p_1 - p_0}{r_0^3 - r_1^3} \; \frac{r_0^3 r_1^3}{r^3},$$

(3) tension in all directions at right angles to the radius vector of amount

$$\tfrac{1}{2}\, \frac{p_1 - p_0}{r_0^3 - r_1^3} \; \frac{r_0^3 r_1^3}{r^3}.$$

The corresponding strains consist of

(1) uniform extension of all lines of the body of amount

$$\frac{1}{3k} \; \frac{p_1 r_1^3 - p_0 r_0^3}{r_0^3 - r_1^3},$$

(2) radial contraction of amount

$$\frac{1}{2\mu} \; \frac{p_1 - p_0}{r_0^3 - r_1^3} \; \frac{r_0^3 r_1^3}{r^3},$$

(3) extension in all directions at right angles to the radius vector of amount

$$\frac{1}{4\mu} \; \frac{p_1 - p_0}{r_0^3 - r_1^3} \; \frac{r_0^3 r_1^3}{r^3},$$

where μ is the modulus of rigidity of the material, $= \tfrac{1}{2}E / (1 + \sigma)$. The volume included between the two surfaces of the body is increased by the fraction $(p_1 r_1^3 - p_0 r_0^3) / k(r_0^3 - r_1^3)$ of itself, and the volume within the inner surface is increased by the fraction

$$\frac{3\,(p_1 - p_0)}{4\mu} \; \frac{r_0^3}{r_0^3 - r_1^3} + \frac{p_1 r_1^3 - p_0 r_0^3}{k\,(r_0^3 - r_1^3)}$$

of itself. For a shell subject only to internal pressure p the greatest extension is the extension at right angles to the radius at the inner surface, and its amount is

$$\frac{p r_1^3}{r_0^3 - r_1^3} \left(\frac{1}{3k} + \frac{1}{4\mu} \; \frac{r_0^3}{r_1^3} \right);$$

the greatest tension is the transverse tension at the inner surface, and its amount is $p\, (\tfrac{1}{2}\, r_0^3 + r_1^3) / (r_0^3 - r_1^3)$.

77. In the problem of a cylindrical shell under pressure a complication may arise from the effects of the ends; but when the ends are free from stress the solution is very simple. With notation similar to that in § 76 it can be shown that the stress at a distance r from the axis consists of

(1) uniform tension in all directions at right angles to the axis of amount

$$\frac{p_1 r_1^2 - p_0 r_0^2}{r_0^2 - r_1^2},$$

(2) radial pressure of amount

$$\frac{p_1 - p_0}{r_0^2 - r_1^2} \; \frac{r_0^2 r_1^2}{r^2},$$

(3) hoop tension numerically equal to this radial pressure.

The corresponding strains consist of

(1) uniform extension of all lines of the material at right angles to the axis of amount

$$\frac{1-\sigma}{E}\ \frac{p_1 r_1^2 - p_0 r_0^2}{r_0^2 - r_1^2},$$

(2) radial contraction of amount

$$\frac{1+\sigma}{E}\ \frac{p_1 - p_0}{r_0^2 - r_1^2}\ \frac{r_0^2 r_1^2}{r^2},$$

(3) extension along the circular filaments numerically equal to this radial contraction,

(4) uniform contraction of the longitudinal filaments of amount

$$\frac{2\sigma}{E}\ \frac{p_1 r_1^2 - p_0 r_0^2}{r_0^2 - r_1^2}.$$

For a shell subject only to internal pressure p the greatest extension is the circumferential extension at the inner surface, and its amount is

$$\frac{p}{E}\left(\frac{r_0^2 + r_1^2}{r_0^2 - r_1^2} + \sigma\right);$$

the greatest tension is the hoop tension at the inner surface, and its amount is $p\,(r_0^2 + r_1^2) / (r_0^2 - r_1^2)$.

78. When the ends of the tube, instead of being free, are closed by disks, so that the tube becomes a closed cylindrical vessel, the longitudinal extension is determined by the condition that the resultant longitudinal tension in the walls balances the resultant normal pressure on either end. This condition gives the value of the extension of the longitudinal filaments as

$$(p_1 r_1^2 - p_0 r_0^2) / 3k\,(r_0^2 - r_1^2),$$

where k is the modulus of compression of the material. The result may be applied to the experimental determination of k, by measuring the increase of length of a tube subjected to internal pressure (A. Mallock, *Proc. R. Soc. London*, lxxiv., 1904, and C. Chree, *ibid.*).

79. The results obtained in § 77 have been applied to gun construction; we may consider that one cylinder is heated so as to slip over another upon which it shrinks by cooling, so that the two form a single body in a condition of initial stress.

We take P as the measure of the pressure between the two, and p for the pressure within the inner cylinder by which the system is afterwards strained, and denote by r′ the radius of the common surface. To obtain the stress at any point we superpose the system consisting of radial pressure $p\,(r_1^2/r^2) \cdot (r_0^2 - r^2) / (r_0^2 - r_1^2)$ and hoop tension $p\,(r_1^2/r^2) \cdot (r_0^2 + r^2) / (r_0^2 - r_1^2)$ upon a system which, for the outer cylinder, consists of radial pressure $P\,(r'^2/r^2) \cdot (r_0^2 - r^2) / (r_0^2 - r'^2)$ and hoop tension $P\,(r'^2/r^2) \cdot (r_0^2 + r^2) / (r_0^2 - r'^2)$, and for the inner cylinder consists of radial pressure $P\,(r'^2/r^2) \cdot (r^2 - r_1^2) / (r'^2 - r_1^2)$ and hoop tension $P\,(r'^2/r^2) \cdot (r^2 + r_1^2) / (r'^2 - r_1^2)$. The hoop tension at the inner surface is less than it would be for a tube of equal thickness without initial stress in the ratio

$$1 - \frac{P}{p}\ \frac{2r'^2}{r_0^2 + r_1^2}\ \frac{r_0^2 + r_1^2}{r'^2 - r_1^2} : 1.$$

This shows how the strength of the tube is increased by the initial stress. When the initial stress is produced by tightly wound wire, a similar gain of strength accrues.

80. In the problem of determining the distribution of stress and strain in a circular cylinder, rotating about its axis, simple solutions have been obtained which are sufficiently exact for the two special cases of a thin disk and a long shaft.

Suppose that a circular disk of radius a and thickness 2l, and of density ρ, rotates about its axis with angular velocity ω, and consider the following systems of superposed stresses at any point distant r from the axis and z from the middle plane:

(1) uniform tension in all directions at right angles to the axis of amount $\frac{1}{8}\,\omega^2\rho a^2\,(3 + \sigma)$,

(2) radial pressure of amount $\frac{1}{8}\,\omega^2\rho r^2\,(3 + \sigma)$,

(3) pressure along the circular filaments of amount $\frac{1}{8}\,\omega^2\rho r^2\,(1 + 3\sigma)$,

(4) uniform tension in all directions at right angles to the axis of amount $\frac{1}{6}\,\omega^2\rho\,(l^2 - 3z^2)\,\sigma\,(1 + \sigma) / (1 - \sigma)$.

The corresponding strains may be expressed as

(1) uniform extension of all filaments at right angles to the axis of amount

$$\frac{1-\sigma}{E}\ \tfrac{1}{8}\,\omega^2\rho a^2\,(3 + \sigma),$$

(2) radial contraction of amount

$$\frac{1-\sigma^2}{E}\ \tfrac{3}{8}\,\omega^2\rho r^2,$$

(3) contraction along the circular filaments of amount

$$\frac{1-\sigma^2}{E}\ \tfrac{1}{8}\,\omega^2\rho r^2,$$

(4) extension of all filaments at right angles to the axis of amount

$$\frac{1}{E} \tfrac{1}{6}\, \omega^2\rho\, (l^2 - 3_z^2)\, \sigma\, (1 + \sigma),$$

(5) contraction of the filaments normal to the plane of the disk of amount

$$\frac{2\sigma}{E} \tfrac{1}{8}\, \omega^2\rho a^2\, (3 + \sigma) - \frac{\sigma}{E} \tfrac{1}{2}\, \omega^2\rho r^2\, (1 + \sigma) + \frac{2\sigma}{E} \tfrac{1}{6}\, \omega^2\rho\, (l^2 - 3z^2)\, \sigma\, \frac{(1 + \sigma)}{(1 - \sigma)}.$$

The greatest extension is the circumferential extension near the centre, and its amount is

$$\frac{(3 + \sigma)\,(1 - \sigma)}{8E}\, \omega^2\rho a^2 + \frac{\sigma\,(1 + \sigma)}{6E}\, \omega^2\rho l^2.$$

The longitudinal contraction is required to make the plane faces of the disk free from pressure, and the terms in l and z enable us to avoid tangential traction on any cylindrical surface. The system of stresses and strains thus expressed satisfies all the conditions, except that there is a small radial tension on the bounding surface of amount per unit area $\tfrac{1}{6}\, \omega^2\rho\, (l^2 - 3z^2)\, \sigma\, (1 + \sigma) / (1 - \sigma)$. The resultant of these tensions on any part of the edge of the disk vanishes, and the stress in question is very small in comparison with the other stresses involved when the disk is thin; we may conclude that, for a thin disk, the expressions given represent the actual condition at all points which are not very close to the edge (cf. § 55). The effect to the longitudinal contraction is that the plane faces become slightly concave (fig. 32).

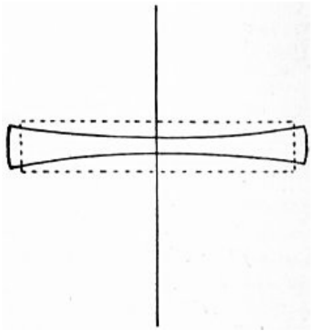

FIG. 32.

81. The corresponding solution for a disk with a circular axle-hole (radius b) will be obtained from that given in the last section by superposing the following system of additional stresses:

(1) radial tension of amount $\tfrac{1}{8}\, \omega^2\rho b^2\, (1 - a^2/r^2)\, (3 + \sigma)$,

(2) tension along the circular filaments of amount

$$\tfrac{1}{8}\, \omega^2\rho b^2\, (1 + a^2/r^2)\, (3 + \sigma).$$

The corresponding additional strains are

(1) radial contraction of amount

$$\frac{3 + \sigma}{8E} \left\{ (1 + \sigma) \frac{a^2}{r^2} - (1 - \sigma) \right\} \omega^2\rho b^2,$$

(2) extension along the circular filaments of amount

$$\frac{3 + \sigma}{8E} \left\{ (1 + \sigma) \frac{a^2}{r^2} + (1 - \sigma) \right\} \omega^2\rho b^2.$$

(3) contraction of the filaments parallel to the axis of amount

$$\frac{\sigma\,(3 + \sigma)}{4E}\, \omega^2\rho b^2.$$

Again, the greatest extension is the circumferential extension at the inner surface, and, when the hole is very small, its amount is nearly double what it would be for a complete disk.

82. In the problem of the rotating shaft we have the following stress-system:

(1) radial tension of amount $\tfrac{1}{8}\, \omega^2\rho\, (a^2 - r^2)\, (3 - 2\sigma) / (1 - \sigma)$,

(2) circumferential tension of amount

$$\tfrac{1}{8}\, \omega^2\rho\, \{a^2\, (3 - 2\sigma) / (1 - \sigma) - r^2\, (1 + 2\sigma) / (1 - \sigma)\},$$

(3) longitudinal tension of amount $\tfrac{1}{4}\, \omega^2\rho\, (a^2 - 2r^2)\, \sigma / (1 - \sigma)$.

The resultant longitudinal tension at any normal section vanishes, and the radial tension vanishes at the bounding surface; and thus the expressions here given may be taken to represent the actual condition at all points which are not very close to the ends of the shaft. The contraction of the longitudinal filaments is uniform and equal to $\tfrac{1}{2}\, \omega^2\rho a^2\sigma / E$. The greatest extension in the rotating shaft is the circumferential extension close to the axis, and its amount is $\tfrac{1}{8}\, \omega^2\rho a^2\, (3 - 5\sigma) / E\, (1 - \sigma)$.

The value of any theory of the strength of long rotating shafts founded on these formulae is diminished by the circumstance that at sufficiently high speeds the shaft may tend to take up a curved form, the straight form being unstable. The shaft is then said to *whirl*. This occurs when the period of rotation of the shaft is very nearly coincident with one of its periods of lateral vibration. The lowest speed at which whirling can take place in a shaft of length l, freely supported at its ends, is given by the formula

$$\omega^2\rho = \tfrac{1}{4}\, Ea^2\, (\pi/l)^4.$$

As in § 61, this formula should not be applied unless the length of the shaft is a considerable multiple of its diameter. It implies that whirling is to be expected whenever ω approaches this critical value.

83. When the forces acting upon a spherical or cylindrical body are not radial, the problem becomes more complicated. In the case of the sphere deformed by any forces it has been completely solved, and the solution has been applied by Lord Kelvin and Sir G.H. Darwin to many interesting questions of cosmical physics. The 158
nature of the stress produced in the interior of the earth by the weight of continents and mountains, the spheroidal figure of a rotating solid planet, the rigidity of the earth, are among the questions which have in this way been attacked. Darwin concluded from his investigation that, to support the weight of the existing continents and mountain ranges, the materials of which the earth is composed must, at great depths (1600 kilometres), have at least the strength of granite. Kelvin concluded from his investigation that the actual

heights of the tides in the existing oceans can be accounted for only on the supposition that the interior of the earth is solid, and of rigidity nearly as great as, if not greater than, that of steel.

84. Some interesting problems relating to the strains produced in a cylinder of finite length by forces distributed symmetrically round the axis have been solved. The most important is that of a cylinder crushed between parallel planes in contact with its plane ends. The solution was applied to explain the discrepancies that have been observed in different tests of crushing strength according as the ends of the test specimen are or are not prevented from spreading. It was applied also to explain the fact that in such tests small conical pieces are sometimes cut out at the ends subjected to pressure.

85. *Vibrations and Waves.*—When a solid body is struck, or otherwise suddenly disturbed, it is thrown into a state of vibration. There always exist dissipative forces which tend to destroy the vibratory motion, one cause of the subsidence of the motion being the communication of energy to surrounding bodies. When these dissipative forces are disregarded, it is found that an elastic solid body is capable of vibrating in such a way that the motion of any particle is simple harmonic motion, all the particles completing their oscillations in the same period and being at any instant in the same phase, and the displacement of any selected one in any particular direction bearing a definite ratio to the displacement of an assigned one in an assigned direction. When a body is moving in this way it is said to be *vibrating in a normal mode*. For example, when a tightly stretched string of negligible flexural rigidity, such as a violin string may be taken to be, is fixed at the ends, and vibrates transversely in a normal mode, the displacements of all the particles have the same direction, and their magnitudes are proportional at any instant to the ordinates of a curve of sines. Every body possesses an infinite number of normal modes of vibration, and the *frequencies* (or numbers of vibrations per second) that belong to the different modes form a sequence of increasing numbers. For the string, above referred to, the fundamental tone and the various overtones form an harmonic scale, that is to say, the frequencies of the normal modes of vibration are proportional to the integers 1, 2, 3, ... In all these modes except the first the string vibrates as if it were divided into a number of equal pieces, each having fixed ends; this number is in each case the integer defining the frequency. In general the normal modes of vibration of a body are distinguished one from another by the number and situation of the surfaces (or other *loci*) at which some characteristic displacement or traction vanishes. The problem of determining the normal modes and frequencies of free vibration of a body of definite size, shape and constitution, is a mathematical problem of a similar character to the problem of determining the state of stress in the body when subjected to given forces. The bodies which have been most studied are strings and thin bars, membranes, thin plates and shells, including bells, spheres and cylinders. Most of the results are of special importance in their bearing upon the theory of sound.

86. The most complete success has attended the efforts of mathematicians to solve the problem of free vibrations for an isotropic sphere. It appears that the modes of vibration fall into two classes: one characterized by the absence of a radial component of displacement, and the other by the absence of a radial component of rotation (§ 14). In each class there is a doubly infinite number of modes. The displacement in any mode is determined in terms of a single spherical harmonic function, so that there are modes of each class corresponding to spherical harmonics of every integral degree; and for each degree there is an infinite number of modes, differing from one another in the number and position of the concentric spherical surfaces at which some characteristic displacement vanishes. The most interesting modes are those in which the sphere becomes slightly spheroidal, being alternately prolate and oblate during the course of a vibration; for these vibrations tend to be set up in a spherical planet by tide-generating forces. In a sphere of the size of the earth, supposed to be incompressible and as rigid as steel, the period of these vibrations is 66 minutes.

87. The theory of free vibrations has an important bearing upon the question of the strength of structures subjected to sudden blows or shocks. The stress and strain developed in a body by sudden applications of force may exceed considerably those which would be produced by a gradual application of the same forces. Hence there arises the general question of *dynamical resistance*, or of the resistance of a body to forces applied so quickly that the inertia of the body comes sensibly into play. In regard to this question we have two chief theoretical results. The first is that the strain produced by a force suddenly applied may be as much as twice the statical strain, that is to say, as the strain which would be produced by the same force when the body is held in equilibrium under its action; the second is that the sudden reversal of the force may produce a strain three times as great as the statical strain. These results point to the importance of specially strengthening the parts of any machine (*e.g.* screw propeller shafts) which are subject to sudden applications or reversals of load. The theoretical limits of twice, or three times, the statical strain are not in general attained. For example, if a thin bar hanging vertically from its upper end is suddenly loaded at its lower end with a weight equal to its own weight, the greatest dynamical strain bears to the greatest statical strain the ratio 1.63 : 1; when the attached weight is four times the weight of the bar the ratio becomes 1.84 : 1. The method by which the result just mentioned is reached has recently been applied to the question of the breaking of winding ropes used in mines. It appeared that, in order to bring the results into harmony with the observed facts, the strain in the supports must be taken into account as well as the strain in the rope (J. Perry, *Phil. Mag.*, 1906 (vi.), vol. ii.).

88. The immediate effect of a blow or shock, locally applied to a body, is the generation of a wave which travels through the body from the locality first affected. The question of the propagation of waves through an elastic solid body is historically of very great importance; for the first really successful efforts to construct a theory of elasticity (those of S.D. Poisson, A.L. Cauchy and G. Green) were prompted, at least in part, by Fresnel's theory of the propagation of light by transverse vibrations. For many years the luminiferous medium was identified with the isotropic solid of the theory of elasticity. Poisson showed that a disturbance communicated to the body gives rise to two waves which are propagated through it with different velocities; and Sir G.G. Stokes afterwards showed that the quicker wave is a wave of irrotational dilatation, and the slower wave is a wave of rotational distortion accompanied by no change of volume. The velocities of the two waves in a solid of density ρ are $\sqrt{\{(\lambda + 2\mu)/\rho\}}$ and $\sqrt{(\mu/\rho)}$, λ and μ being the constants so denoted in § 26. When the surface of the body is free from traction, the waves on reaching the surface are reflected; and thus after a little time the body would, if there were no dissipative forces, be in a very complex state of motion due to multitudes of waves passing to and fro through it. This state can be expressed as a state of vibration, in

which the motions belonging to the various normal modes (§ 85) are superposed, each with an appropriate amplitude and phase. The waves of dilatation and distortion do not, however, give rise to different modes of vibration, as was at one time supposed, but any mode of vibration in general involves both dilatation and rotation. There are exceptional results for solids of revolution; such solids possess normal modes of vibration which involve no dilatation. The existence of a boundary to the solid body has another effect, besides reflexion, upon the propagation of waves. Lord Rayleigh has shown that any disturbance originating at the surface gives rise to waves which travel away over the surface as well as to waves which travel through the interior; and any internal disturbance, on reaching the surface, also gives rise to such superficial waves. The velocity of the superficial waves is a little less than that of the waves of distortion: $0.9554 \sqrt{(\mu/\rho)}$ when the 159
material is incompressible $0.9194 \sqrt{(\mu/\rho)}$ when the Poisson’s ratio belonging to the material is ¼.

89. These results have an application to the propagation of earthquake shocks (see also Earthquake). An internal disturbance should, if the earth can be regarded as solid, give rise to three wave-motions: two propagated through the interior of the earth with different velocities, and a third propagated over the surface. The results of seismographic observations have independently led to the recognition of three phases of the recorded vibrations: a set of “preliminary tremors” which are received at different stations at such times as to show that they are transmitted directly through the interior of the earth with a velocity of about 10 km. per second, a second set of preliminary tremors which are received at different stations at such times as to show that they are transmitted directly through the earth with a velocity of about 5 km. per second, and a “main shock,” or set of large vibrations, which becomes sensible at different stations at such times as to show that a wave is transmitted over the surface of the earth with a velocity of about 3 km. per second. These results can be interpreted if we assume that the earth is a solid body the greater part of which is practically homogeneous, with high values for the rigidity and the resistance to compression, while the superficial portions have lower values for these quantities. The rigidity of the central portion would be about $(1.4)10^{12}$ dynes per square cm., which is considerably greater than that of steel, and the resistance to compression would be about $(3.8)10^{12}$ dynes per square cm. which is much greater than that of any known material. The high value of the resistance to compression is not surprising when account is taken of the great pressures, due to gravitation, which must exist in the interior of the earth. The high value of the rigidity can be regarded as a confirmation of Lord Kelvin’s estimate founded on tidal observations (§ 83).

90. *Strain produced by Heat.*—The mathematical theory of elasticity as at present developed takes no account of the strain which is produced in a body by unequal heating. It appears to be impossible in the present state of knowledge to form as in § 39 a system of differential equations to determine both the stress and the temperature at any point of a solid body the temperature of which is liable to variation. In the cases of isothermal and adiabatic changes, that is to say, when the body is slowly strained without variation of temperature, and also when the changes are effected so rapidly that there is no gain or loss of heat by any element, the internal energy of the body is sufficiently expressed by the strain-energy-function (§§ 27, 30). Thus states of equilibrium and of rapid vibration can be determined by the theory that has been explained above. In regard to thermal effects we can obtain some indications from general thermodynamic theory. The following passages extracted from the article “Elasticity” contributed to the 9th edition of the *Encyclopaedia Britannica* by Sir W. Thomson (Lord Kelvin) illustrate the nature of these indications:—“From thermodynamic theory it is concluded that cold is produced whenever a solid is strained by opposing, and heat when it is strained by yielding to, any elastic force of its own, the strength of which would diminish if the temperature were raised; but that, on the contrary, heat is produced when a solid is strained against, and cold when it is strained by yielding to, any elastic force of its own, the strength of which would increase if the temperature were raised. When the strain is a condensation or dilatation, uniform in all directions, a fluid may be included in the statement. Hence the following propositions:—

“(1) A cubical compression of any elastic fluid or solid in an ordinary condition causes an evolution of heat; but, on the contrary, a cubical compression produces cold in any substance, solid or fluid, in such an abnormal state that it would contract if heated while kept under constant pressure. Water below its temperature (3.9° Cent.) of maximum density is a familiar instance.

“(2) If a wire already twisted be suddenly twisted further, always, however, within its limits of elasticity, cold will be produced; and if it be allowed suddenly to untwist, heat will be evolved from itself (besides heat generated externally by any work allowed to be wasted, which it does in untwisting). It is assumed that the torsional rigidity of the wire is diminished by an elevation of temperature, as the writer of this article had found it to be for copper, iron, platinum and other metals.

“(3) A spiral spring suddenly drawn out will become lower in temperature, and will rise in temperature when suddenly allowed to draw in. [This result has been experimentally verified by Joule (’Thermodynamic Properties of Solids,’ *Phil. Trans.*, 1858) and the amount of the effect found to agree with that calculated, according to the preceding thermodynamic theory, from the amount of the weakening of the spring which he found by experiment.]

“(4) A bar or rod or wire of any substance with or without a weight hung on it, or experiencing any degree of end thrust, to begin with, becomes cooled if suddenly elongated by end pull or by diminution of end thrust, and warmed if suddenly shortened by end thrust or by diminution of end pull; except abnormal cases in which with constant end pull or end thrust elevation of temperature produces shortening; in every such case pull or diminished thrust produces elevation of temperature, thrust or diminished pull lowering of temperature.

“(5) An india-rubber band suddenly drawn out (within its limits of elasticity) becomes warmer; and when allowed to contract, it becomes colder. Any one may easily verify this curious property by placing an india-rubber band in slight contact with the edges of the lips, then suddenly extending it—it becomes very perceptibly warmer: hold it for some time stretched nearly to breaking, and then suddenly allow it to shrink—it becomes quite startlingly colder, the cooling effect being sensible not merely to the lips but to the fingers holding the band. The first published statement of this curious observation is due to J. Gough (*Mem. Lit. Phil. Soc. Manchester*, 2nd series, vol. i. p. 288), quoted by Joule in his paper on ‘Thermodynamic Properties of

Solids’ (cited above). The thermodynamic conclusion from it is that an india-rubber band, stretched by a constant weight of sufficient amount hung on it, must, when heated, pull up the weight, and, when cooled, allow the weight to descend: this Gough, independently of thermodynamic theory, had found to be actually the case. The experiment any one can make with the greatest ease by hanging a few pounds weight on a common india-rubber band, and taking a red-hot coal in a pair of tongs, or a red-hot poker, and moving it up and down close to the band. The way in which the weight rises when the red-hot body is near, and falls when it is removed, is quite startling. Joule experimented on the amount of shrinking per degree of elevation of temperature, with different weights hung on a band of vulcanized india-rubber, and found that they closely agreed with the amounts calculated by Thomson’s theory from the heating effects of pull, and cooling effects of ceasing to pull, which he had observed in the same piece of india-rubber.”

91. *Initial Stress.*—It has been pointed out above (§ 20) that the “unstressed” state, which serves as a zero of reckoning for strains and stresses is never actually attained, although the strain (measured from this state), which exists in a body to be subjected to experiment, may be very slight. This is the case when the “initial stress,” or the stress existing before the experiment, is small in comparison with the stress developed during the experiment, and the limit of linear elasticity (§ 32) is not exceeded. The existence of initial stress has been correlated above with the existence of body forces such as the force of gravity, but it is not necessarily dependent upon such forces. A sheet of metal rolled into a cylinder, and soldered to maintain the tubular shape, must be in a state of considerable initial stress quite apart from the action of gravity. Initial stress is utilized in many manufacturing processes, as, for example, in the construction of ordnance, referred to in § 79, in the winding of golf balls by means of india-rubber in a state of high tension (see the report of the case *The Haskell Golf Ball Company* v. *Hutchinson & Main* in *The Times* of March 1, 1906). In the case of a body of ordinary dimensions it is such internal stress as this which is especially meant by the phrase “initial 160 stress.” Such a body, when in such a state of internal stress, is sometimes described as “self-strained.” It would be better described as “self-stressed.” The somewhat anomalous behaviour of cast iron has been supposed to be due to the existence within the metal of initial stress. As the metal cools, the outer layers cool more rapidly than the inner, and thus the state of initial stress is produced. When cast iron is tested for tensile strength, it shows at first no sensible range either of perfect elasticity or of linear elasticity; but after it has been loaded and unloaded several times its behaviour begins to be more nearly like that of wrought iron or steel. The first tests probably diminish the initial stress.

92. From a mathematical point of view the existence of initial stress in a body which is “self-stressed” arises from the fact that the equations of equilibrium of a body free from body forces or surface tractions, viz. the equations of the type

$$\frac{\partial X_x}{\partial x} + \frac{\partial X_y}{\partial y} + \frac{\partial Z_x}{\partial z} = 0,$$

possess solutions which differ from zero. If, in fact, φ_1, φ_2, φ_3 denote any arbitrary functions of *x, y, z*, the equations are satisfied by putting

$$X_x = \frac{\partial^2 \varphi_3}{\partial y^2} + \frac{\partial^2 \varphi_2}{\partial z}, \ldots, Y_z = -\frac{\partial^2 \varphi_1}{\partial y \partial z}, \ldots;$$

and it is clear that the functions φ_1, φ_2, φ_3 can be adjusted in an infinite number of ways so that the bounding surface of the body may be free from traction.

93. Initial stress due to body forces becomes most important in the case of a gravitating planet. Within the earth the stress that arises from the mutual gravitation of the parts is very great. If we assumed the earth to be an elastic solid body with moduluses of elasticity no greater than those of steel, the strain (measured from the unstressed state) which would correspond to the stress would be much too great to be calculated by the ordinary methods of the theory of elasticity (§ 75). We require therefore some other method of taking account of the initial stress. In many investigations, for example those of Lord Kelvin and Sir G.H. Darwin referred to in § 83, the difficulty is turned by assuming that the material may be treated as practically incompressible; but such investigations are to some extent incomplete, so long as the corrections due to a finite, even though high, resistance to compression remain unknown. In other investigations, such as those relating to the propagation of earthquake shocks and to gravitational instability, the possibility of compression is an essential element of the problem. By gravitational instability is meant the tendency of gravitating matter to condense into nuclei when slightly disturbed from a state of uniform diffusion; this tendency has been shown by J.H. Jeans (*Phil. Trans*. A. 201, 1903) to have exerted an important influence upon the course of evolution of the solar system. For the treatment of such questions Lord Rayleigh (*Proc. R. Soc. London*, A. 77, 1906) has advocated a method which amounts to assuming that the initial stress is hydrostatic pressure, and that the actual state of stress is to be obtained by superposing upon this initial stress a stress related to the state of strain (measured from the initial state) by the same formulae as hold for an elastic solid body free from initial stress. The development of this method is likely to lead to results of great interest.

Authorities.—In regard to the analysis requisite to prove the results set forth above, reference may be made to A.E.H. Love, *Treatise on the Mathematical Theory of Elasticity* (2nd ed., Cambridge, 1906), where citations of the original authorities will also be found. The following treatises may be mentioned: Navier, *Résumé des leçons sur l’application de la mécanique* (3rd ed., with notes by Saint-Venant, Paris, 1864); G. Lamé, *Leçons sur la théorie mathématique de l’élasticité des corps solides* (Paris, 1852); A. Clebsch, *Theorie der Elasticität fester Körper* (Leipzig, 1862; French translation with notes by Saint-Venant, Paris, 1883); F. Neumann, *Vorlesungen über die Theorie der Elasticität* (Leipzig, 1885); Thomson and Tait, *Natural Philosophy* (Cambridge, 1879, 1883); Todhunter and Pearson, *History of the Elasticity and Strength of Materials* (Cambridge, 1886-1893). The article “Elasticity” by Sir W. Thomson (Lord Kelvin) in 9th ed. of *Encyc. Brit.* (reprinted in his *Mathematical and Physical Papers*, iii., Cambridge, 1890) is especially valuable, not only for the exposition of the theory and its practical applications, but also for the tables of physical constants which are there given.

(A. E. H. L.)

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1 The sign of M is shown by the arrow-heads in fig. 19, for which, with y downwards,

$$EI \frac{d^2y}{dx^2} + M = 0.$$

2 The figure is drawn for a case where the bending moment has the same sign throughout.

3 M_0 is taken to have, as it obviously has, the opposite sense to that shown in fig. 19.

4 The line joining the points of contact must be normal to the planes.

ELATERITE, also termed Elastic Bitumen and Mineral Caoutchouc, a mineral hydrocarbon, which occurs at Castleton in Derbyshire, in the lead mines of Odin and elsewhere. It varies somewhat in consistency, being sometimes soft, elastic and sticky; often closely resembling india-rubber; and occasionally hard and brittle. It is usually dark brown in colour and slightly translucent. A substance of similar physical character is found in the Coorong district of South Australia, and is hence termed coorongite, but Prof. Ralph Tate considers this to be a vegetable product.

ELATERIUM, a drug consisting of a sediment deposited by the juice of the fruit of *Ecballium Elaterium*, the squirting cucumber, a native of the Mediterranean region. The plant, which is a member of the natural order Cucurbitaceae, resembles the vegetable marrow in its growth. The fruit resembles a small cucumber, and when ripe is highly turgid, and separates almost at a touch from the fruit stalk. The end of the stalk forms a stopper, on the removal of which the fluid contents of the fruit, together with the seeds, are squirted through the aperture by the sudden contraction of the wall of the fruit. To prepare the drug the fruit is sliced lengthwise and slightly pressed; the greenish and slightly turbid juice thus obtained is strained and set aside; and the deposit of elaterium formed after a few hours is collected on a linen filter, rapidly drained, and dried on porous tiles at a gentle heat. Elaterium is met with in commerce in light, thin, friable, flat or slightly incurved opaque cakes, of a greyish-green colour, bitter taste and tea-like smell.

The drug is soluble in alcohol, but insoluble in water and ether. The official dose is $\frac{1}{10}$-$\frac{1}{2}$ grain, and the British pharmacopeia directs that the drug is to contain from 20 to 25% of the active principle elaterinum or elaterin. A resin in the natural product aids its action. Elaterin is extracted from elaterium by chloroform and then precipitated by ether. It has the formula $C_{20}H_{28}O_5$. It forms colourless scales which have a bitter taste, but it is highly inadvisable to taste either this substance or elaterium. Its dose is $\frac{1}{40}$-$\frac{1}{10}$ grain, and the British pharmacopeia contains a useful preparation, the Pulvis Elaterini Compositus, which contains one part of the active principle in forty.

The action of this drug resembles that of the saline aperients, but is much more powerful. It is the most active hydragogue purgative known, causing also much depression and violent griping. When injected subcutaneously it is inert, as its action is entirely dependent upon its admixture with the bile. The drug is undoubtedly valuable in cases of dropsy and Bright’s disease, and also in cases of cerebral haemorrhage, threatened or present. It must not be used except in urgent cases, and must invariably be employed with the utmost care, especially if the state of the heart be unsatisfactory.

ELBA (Gr. Αἰθαλία; Lat. *Ilva*), an island off the W. coast of Italy, belonging to the province of Leghorn, from which it is 45 m. S., and 7 m. S.W. of Piombino, the nearest point of the mainland. Pop. (1901) 25,043 (including Pianosa). It is about 19 m. long, 6½ m. broad, and 140 sq. m. in area; and its highest point is 3340 ft. (Monte Capanne). It forms, like Giglio and Monte Cristo, part of a sunken mountain range extending towards Corsica and Sardinia.

The oldest rocks of Elba consist of schist and serpentine which in the eastern part of the island are overlaid by beds containing Silurian and Devonian fossils. The Permian may be represented, but the Trias is absent, and in general the older Palaeozoic rocks are overlaid directly by the Rhaetic and Lias. The Liassic beds are often metamorphosed and the limestones contain garnet and wollastonite. The next geological formation which is represented is the Eocene, consisting of nummulitic limestone, sandstone and schist. The Miocene and Pliocene are absent. The most remarkable feature in the geology of Elba is the extent of the granitic and ophiolitic eruptions of the Tertiary period. Serpentines, peridotites and diabases are interstratified with the Eocene deposits. The granite, which is intruded through the Eocene beds, is associated with a pegmatite containing tourmaline and cassiterite. The celebrated iron ore of Elba is of Tertiary age and occurs indifferently in all the older rocks. The deposits are superficial, resulting from the opening out of veins at the surface, and consist chiefly of haematite. These ores were worked by the ancients, but so inefficiently that their spoil-heaps can be smelted again with profit. This process is now gone through on the island itself. The granite was also quarried by the Romans, but is not now much worked. 161

Parts of the island are fertile, and the cultivation of vines, and the tunny and sardine fishery, also give employment to a part of the population. The capital of the island is Portoferraio—pop. (1901) 5987—in the centre of the N. coast, enclosed by an amphitheatre of lofty mountains, the slopes of which are covered with villas and gardens. This is the best harbour, the ancient *Portus Argous*. The town was built and fortified by Cosimo I. in 1548, who called it Cosmopolis. Above the harbour, between the forts Stella and Falcone, is the palace of Napoleon I., and 4 m. to the S.W. is his villa; while on the N. slope of Monte Capanne is another of his country houses. The other villages in the island are Campo nell' Elba, on the S. near the W. end, Marciana and Marciana Marina on the N. of the island near the W. extremity, Porto Longone, on the E. coast, with picturesque Spanish fortifications, constructed in 1602 by Philip III.; Rio dell' Elba and Rio Marina, both on the E. side of the island, in the mining district. At Le Grotte, between Portoferraio and Rio dell' Elba, and at Capo Castello, on the N.E. of the island, are ruins of Roman date.

Elba was famous for its mines in early times, and the smelting furnaces gave it its Greek name of Α᾿ θαλία ("soot island"). In Roman times, and until 1900, however, owing to lack of fuel, the smelting was done on the mainland. In 453 B.C. Elba was devastated by a Syracusan squadron. From the 11th to the 14th century it belonged to Pisa, and in 1399 came under the dukes of Piombino. In 1548 it was ceded by them to Cosimo I. of Florence. In 1596 Porto Longone was taken by Philip III. of Spain, and retained until 1709, when it was ceded to Naples. In 1802 the island was given to France by the peace of Amiens. On Napoleon's deposition, the island was ceded to him with full sovereign rights, and he resided there from the 5th of May 1814 to the 26th of February 1815. After his fall it was restored to Tuscany, and passed with it to Italy in 1860.

See Sir R. Colt Hoare, *A Tour through the Island of Elba* (London, 1814).

ELBE (the *Albis* of the Romans and the *Labe* of the Czechs), a river of Germany, which rises in Bohemia not far from the frontiers of Silesia, on the southern side of the Riesengebirge, at an altitude of about 4600 ft. Of the numerous small streams (Seifen or Flessen as they are named in the district) whose confluent waters compose the infant river, the most important are the Weisswasser, or White Water, and the Elbseifen, which is formed in the same neighbourhood, but at a little lower elevation. After plunging down the 140 ft. of the Elbfall, the latter stream unites with the steep torrential Weisswasser at Mädelstegbaude, at an altitude of 2230 ft., and thereafter the united stream of the Elbe pursues a southerly course, emerging from the mountain glens at Hohenelbe (1495 ft.), and continuing on at a soberer pace to Pardubitz, where it turns sharply to the west, and at Kolin (730 ft.), some 27 m. farther on, bends gradually towards the north-west. A little above Brandeis it picks up the Iser, which, like itself, comes down from the Riesengebirge, and at Melnik it has its stream more than doubled in volume by the Moldau, a river which winds northwards through the heart of Bohemia in a sinuous, trough-like channel carved through the plateaux. Some miles lower down, at Leitmeritz (433 ft.), the waters of the Elbe are tinted by the reddish Eger, a stream which drains the southern slopes of the Erzgebirge. Thus augmented, and swollen into a stream 140 yds. wide, the Elbe carves a path through the basaltic mass of the Mittelgebirge, churning its way through a deep, narrow rocky gorge. Then the river winds through the fantastically sculptured sandstone mountains of the "Saxon Switzerland," washing successively the feet of the lofty Lilienstein (932 ft. above the Elbe), the scene of one of Frederick the Great's military exploits in the Seven Years' War, Königstein (797 ft. above the Elbe), where in times of war Saxony has more than once stored her national purse for security, and the pinnacled rocky wall of the Bastei, towering 650 ft. above the surface of the stream. Shortly after crossing the Bohemian-Saxon frontier, and whilst still struggling through the sandstone defiles, the stream assumes a north-westerly direction, which on the whole it preserves right away to the North Sea. At Pirna the Elbe leaves behind it the stress and turmoil of the Saxon Switzerland, rolls through Dresden, with its noble river terraces, and finally, beyond Meissen, enters on its long journey across the North German plain, touching Torgau, Wittenberg, Magdeburg, Wittenberge, Hamburg, Harburg and Altona on the way, and gathering into itself the waters of the Mulde and Saale from the left, and those of the Schwarze Elster, Havel and Elde from the right. Eight miles above Hamburg the stream divides into the Norder (or Hamburg) Elbe and the Süder (or Harburg) Elbe, which are linked together by several cross-channels, and embrace in their arms the large island of Wilhelmsburg and some smaller ones. But by the time the river reaches Blankenese, 7 m. below Hamburg, all these anastomosing branches have been reunited, and the Elbe, with a width of 4 to 9 m. between bank and bank, travels on between the green marshes of Holstein and Hanover until it becomes merged in the North Sea off Cuxhaven. At Kolin the width is about 100 ft., at the mouth of the Moldau about 300, at Dresden 960, and at Magdeburg over 1000. From Dresden to the sea the river has a total fall of only 280 ft., although the distance is about 430 m. For the 75 m. between Hamburg and the sea the fall is only 3¼ ft. One consequence of this is that the bed of the river just below Hamburg is obstructed by a bar, and still lower down is choked with sandbanks, so that navigation is confined to a relatively narrow channel down the middle of the stream. But unremitting efforts have been made to maintain a sufficient fairway up to Hamburg (*q.v.*). The tide advances as far as Geesthacht, a little more than 100 m. from the sea. The river is navigable as far as Melnik, that is, the confluence of the Moldau, a distance of 525 m., of which 67 are in Bohemia. Its total length is 725 m., of which 190 are in Bohemia, 77 in the kingdom of Saxony, and 350 in Prussia, the remaining 108 being in Hamburg and other states of Germany. The area of the drainage basin is estimated at 56,000 sq. m.

Navigation.—Since 1842, but more especially since 1871, improvements have been made in the navigability of the Elbe by all the states which border upon its banks. As a result of these labours there is now in the Bohemian portion of the river a minimum depth of 2 ft. 8 in., whilst from the Bohemian frontier down to Magdeburg the minimum depth is 3 ft., and from Magdeburg to Hamburg, 3 ft. 10 in. In 1896 and 1897 Prussia and Hamburg signed covenants whereby two channels are to be kept open to a depth of 9¾ ft., a width of 656 ft., and a length of 550 yds. between Bunthaus and Ortkathen, just above the bifurcation of the Norder Elbe and the Süder Elbe. In 1869 the maximum burden of the vessels which were able to ply on the upper Elbe was 250 tons; but in 1899 it was increased to 800 tons. The large towns through which the river flows have vied with one another in building harbours, providing shipping accommodation, and furnishing

other facilities for the efficient navigation of the Elbe. In this respect the greatest efforts have naturally been made by Hamburg; but Magdeburg, Dresden, Meissen, Riesa, Tetschen, Aussig and other places have all done their relative shares, Magdeburg, for instance, providing a commercial harbour and a winter harbour. In spite, however, of all that has been done, the Elbe remains subject to serious inundations at periodic intervals. Among the worst floods were those of the years 1774, 1799, 1815, 1830, 1845, 1862, 1890 and 1909. The growth of traffic up and down the Elbe has of late years become very considerable. A towing chain, laid in the bed of the river, extends from Hamburg to Aussig, and by this means, as by paddle-tug haulage, large barges are brought from the port of Hamburg into the heart of Bohemia. The fleet of steamers and barges navigating the Elbe is in point of fact greater than on any other German river. In addition to goods thus conveyed, enormous quantities of timber are floated down the Elbe; the weight of the rafts passing the 162
station of Schandau on the Saxon Bohemian frontier amounting in 1901 to 333,000 tons.

A vast amount of traffic is directed to Berlin, by means of the Havel-Spree system of canals, to the Thuringian states and the Prussian province of Saxony, to the kingdom of Saxony and Bohemia, and to the various riverine states and provinces of the lower and middle Elbe. The passenger traffic, which is in the hands of the Sächsisch-Böhmische Dampfschifffahrtsgesellschaft is limited to Bohemia and Saxony, steamers plying up and down the stream from Dresden to Melnik, occasionally continuing the journey up the Moldau to Prague, and down the river as far as Riesa, near the northern frontier of Saxony, and on the average 1½ million passengers are conveyed.

In 1877-1879, and again in 1888-1895, some 100 m. of canal were dug, 5 to 6½ ft. deep and of various widths, for the purpose of connecting the Elbe, through the Havel and the Spree, with the system of the Oder. The most noteworthy of these connexions are the Elbe Canal (14¼ m. long), the Reek Canal (9½ m.), the Rüdersdorfer Gewässer (11½ m.), the Rheinsberger Canal (11¼ m.), and the Sacrow-Paretzer Canal (10 m.), besides which the Spree has been canalized for a distance of 28 m., and the Elbe for a distance of 70 m. Since 1896 great improvements have been made in the Moldau and the Bohemian Elbe, with the view of facilitating communication between Prague and the middle of Bohemia generally on the one hand, and the middle and lower reaches of the Elbe on the other. In the year named a special commission was appointed for the regulation of the Moldau and Elbe between Prague and Aussig, at a cost estimated at about £1,000,000, of which sum two-thirds were to be borne by the Austrian empire and one-third by the kingdom of Bohemia. The regulation is effected by locks and movable dams, the latter so designed that in times of flood or frost they can be dropped flat on the bottom of the river. In 1901 the Austrian government laid before the Reichsrat a canal bill, with proposals for works estimated to take twenty years to complete, and including the construction of a canal between the Oder, starting at Prerau, and the upper Elbe at Pardubitz, and for the canalization of the Elbe from Pardubitz to Melnik (see Austria: *Waterways*). In 1900 Lübeck was put into direct communication with the Elbe at Lauenburg by the opening of the Elbe-Trave Canal, 42 m. in length, and constructed at a cost of £1,177,700, of which the state of Lübeck contributed £802,700, and the kingdom of Prussia £375,000. The canal has been made 72 ft. wide at the bottom, 105 to 126 ft. wide at the top, has a minimum depth of 8⅙ ft., and is equipped with seven locks, each 262½ ft. long and 39¼ ft. wide. It is thus able to accommodate vessels up to 800 tons burden; and the passage from Lübeck to Lauenburg occupies 18 to 21 hours. In the first year of its being open (June 1900 to June 1901) a total of 115,000 tons passed through the canal.[1] A gigantic project has also been put forward for providing water communication between the Rhine and the Elbe, and so with the Oder, through the heart of Germany. This scheme is known as the Midland Canal. Another canal has been projected for connecting Kiel with the Elbe by means of a canal trained through the Plön Lakes.

Bridges.—The Elbe is crossed by numerous bridges, as at Königgrätz, Pardubitz, Kolin, Leitmeritz, Tetschen, Schandau, Pirna, Dresden, Meissen, Torgau, Wittenberg, Rosslau, Barby, Magdeburg, Rathenow, Wittenberge, Dömitz, Lauenburg, and Hamburg and Harburg. At all these places there are railway bridges, and nearly all, but more especially those in Bohemia, Saxony and the middle course of the river—these last on the main lines between Berlin and the west and south-west of the empire—possess a greater or less strategic value. At Leitmeritz there is an iron trellis bridge, 600 yds long. Dresden has four bridges, and there is a fifth bridge at Loschwitz, about 3 m. above the city. Meissen has a railway bridge, in addition to an old road bridge. Magdeburg is one of the most important railway centres in northern Germany; and the Elbe, besides being bridged—it divides there into three arms—several times for vehicular traffic, is also spanned by two fine railway bridges. At both Hamburg and Harburg, again, there are handsome railway bridges, the one (1868-1873 and 1894) crossing the northern Elbe, and the other (1900) the southern Elbe; and the former arm is also crossed by a fine triple-arched bridge (1888) for vehicular traffic.

Fish.—The river is well stocked with fish, both salt-water and fresh-water species being found in its waters, and several varieties of fresh-water fish in its tributaries. The kinds of greatest economic value are sturgeon, shad, salmon, lampreys, eels, pike and whiting.

Tolls.—In the days of the old German empire no fewer than thirty-five different tolls were levied between Melnik and Hamburg, to say nothing of the special dues and privileged exactions of various riparian owners and political authorities. After these had been *de facto*, though not *de jure*, in abeyance during the period of the Napoleonic wars, a commission of the various Elbe states met and drew up a scheme for their regulation, and the scheme, embodied in the Elbe Navigation Acts, came into force in 1822. By this a definite number of tolls, at fixed rates, was substituted for the often arbitrary tolls which had been exacted previously. Still further relief was afforded in 1844 and in 1850, on the latter occasion by the abolition of all tolls between Melnik and the Saxon frontier. But the number of tolls was only reduced to one, levied at Wittenberge, in 1863, about one year after Hanover was induced to give up the Stade or Brunsbüttel toll in return for a compensation of 2,857,340 thalers. Finally, in 1870, 1,000,000 thalers were paid to Mecklenburg and 85,000 thalers to Anhalt, which thereupon abandoned all claims to levy tolls upon the Elbe shipping, and thus navigation on the river became at last entirely free.

History.—The Elbe cannot rival the Rhine in the picturesqueness of the scenery it travels through, nor in the glamour which its romantic and legendary associations exercise over the imagination. But it possesses much to charm the eye in the deep glens of the Riesengebirge, amid which its sources spring, and in the bizarre rock-carving of the Saxon Switzerland. It has been indirectly or directly associated with many stirring events in the history of the German peoples. In its lower course, whatever is worthy of record clusters round the historical vicissitudes of Hamburg—its early prominence as a missionary centre (Ansgar) and as a bulwark against Slav and marauding Northman, its commercial prosperity as a leading member of the Hanseatic League, and its sufferings during the Napoleonic wars, especially at the hands of the ruthless Davoût. The bridge over the river at Dessau recalls the hot assaults of the *condottiere* Ernst von Mansfeld in April 1626, and his repulse by the crafty generalship of Wallenstein. But three years later this imperious leader was checked by the heroic resistance of the "Maiden" fortress of Magdeburg; though two years later still she lost her reputation, and suffered unspeakable horrors at the hands of Tilly's lawless and unlicensed soldiery. Mühlberg, just outside the Saxon frontier, is the place where Charles V. asserted his imperial authority over the Protestant elector of Saxony, John Frederick, the Magnanimous or Unfortunate, in 1547. Dresden, Aussig and Leitmeritz are all reminiscent of the fierce battles of the Hussite wars, and the last named of the Thirty Years' War. But the chief historical associations of the upper (*i.e.* the Saxon and Bohemian) Elbe are those which belong to the Seven Years' War, and the struggle of the great Frederick of Prussia against the power of Austria and her allies. At Pirna (and Lilienstein) in 1756 he caught the entire Saxon army in his fowler's net, after driving back at Lobositz the Austrian forces which were hastening to their assistance; but only nine months later he lost his reputation for "invincibility" by his crushing defeat at Kolin, where the great highway from Vienna to Dresden crosses the Elbe. Not many miles distant, higher up the stream, another decisive battle was fought between the same national antagonists, but with a contrary result, on the memorable 3rd of July 1866.

See M. Buchheister, "Die Elbe u. der Hafen von Hamburg," in *Mitteil. d. Geog. Gesellsch. in Hamburg* (1899), vol. xv. pp. 131-188; V. Kurs, "Die künstlichen Wasserstrassen des deutschen Reichs," in *Geog. Zeitschrift* (1898), pp. 601-617; and (the official) *Der Elbstrom* (1900); B. Weissenborn, *Die Elbzölle und Elbstapelplätze im Mittelalter* (Halle, 1900); Daniel, *Deutschland*; and A. Supan, *Wasserstrassen und Binnenschifffahrt* (Berlin, 1902). 163

1 See *Der Bau des Elbe-Trave Canals und seine Vorgeschichte* (Lübeck, 1900).

ELBERFELD, a manufacturing town of Germany, in the Prussian Rhine province, on the Wupper, and immediately west of and contiguous to Barmen (*q.v.*). Pop. (1816) 21,710; (1840) 31,514; (1885) 109,218; (1905) 167,382. Elberfeld-Barmen, although administratively separate, practically form a single whole. It winds, a continuous strip of houses and factories, for 9 m. along the deep valley, on both banks of the Wupper, which is crossed by numerous bridges, the engirdling hills crowned with woods. Local intercommunication is provided by an electric tramway line and a novel hanging railway—on the Langen mono-rail system—suspended over the bed of the river, with frequent stations. In the centre of the town are a number of irregular and narrow streets, and the river, polluted by the refuse of dye-works and factories, constitutes a constant eyesore. Yet within recent years great alterations have been effected; in the newer quarters are several handsome streets and public buildings; in the centre many insanitary dwellings have been swept away, and their place occupied by imposing blocks of shops and business premises, and a magnificent new town-hall, erected in a dominant position. Among the most recent improvements must be mentioned the Brausenwerther Platz, flanked by the theatre, the public baths, and the railway station and administrative offices. There are eleven Evangelical and five Roman Catholic churches (noticeable among the latter the Suitbertuskirche), a synagogue, and chapels of various other sects. Among other public buildings may be enumerated the civic hall, the law courts and the old town-hall.

The town is particularly rich in educational, industrial, philanthropic and religious institutions. The schools include the Gymnasium (founded in 1592 by the Protestant community as a Latin school), the Realgymnasium (founded in 1830, for "modern" subjects and Latin), the Oberrealschule and Realschule (founded 1893, the latter wholly "modern"), two girls' high schools, a girls' middle-class school, a large number of popular schools, a mechanics' and polytechnic school, a school of mechanics, an industrial drawing school, a commercial school, and a school for the deaf and dumb. There are also a theatre, an institute of music, a library, a museum, a zoological garden, and numerous scientific societies. The town is the seat of the Berg Bible Society. The majority of the inhabitants are Protestant, with a strong tendency towards Pietism; but the Roman Catholics number upwards of 40,000, forming about one-fourth of the total population. The industries of Elberfeld are on a scale of great magnitude. It is the chief centre in Germany of the cotton, wool, silk and velvet manufactures, and of upholstery, drapery and haberdashery of all descriptions, of printed calicoes, of Turkey-red and other dyes, and of fine chemicals. Leather and rubber goods, gold, silver and aluminium wares, machinery, wall-paper, and stained glass are also among other of its staple products. Commerce is lively and the exports to foreign countries are very considerable. The railway system is well devised to meet the requirements of its rapidly increasing trade. Two main lines of railway traverse the valley; that on the south is the main line from Aix-la-Chapelle, Cologne and Düsseldorf to central Germany and Berlin, that on the north feeds the important towns of the Ruhr valley.

The surroundings of Elberfeld are attractive, and public grounds and walks have been recently opened on the hills around with results eminently beneficial to the health of the population.

In the 12th century the site of Elberfeld was occupied by the castle of the lords of Elverfeld, feudatories of the archbishops of Cologne. The fief passed later into the possession of the counts of Berg. The industrial development of the place started with a colony of bleachers, attracted by the clear waters of the Wupper, who in 1532 were granted the exclusive privilege of bleaching yarn. It was not, however, until 1610 that Elberfeld was raised to the status of a town, and in 1640 was surrounded with walls. In 1760 the manufacture of silk was introduced, and dyeing with Turkey-red in 1780; but it was not till the end of the century that its industries developed into importance under the influence of Napoleon’s continental system, which barred out British competition. In 1815 Elberfeld was assigned by the congress of Vienna, with the grand-duchy of Berg, to Prussia, and its prosperity rapidly developed under the Prussian Zollverein.

See Coutelle, *Elberfeld, topographisch-statistische Darstellung* (Elberfeld, 1853); Schell, *Geschichte der Stadt Elberfeld* (1900); A. Shadwell, *Industrial Efficiency* (London, 1906); and Jorde, *Führer durch Elberfeld und seine Umgebung* (1902).

ELBEUF, a town of northern France in the department of Seine-Inférieure, 14 m. S.S.W. of Rouen by the western railway. Pop. (1906) 17,800. Elbeuf, a town of wide, clean streets, with handsome houses and factories, stands on the left bank of the Seine at the foot of hills over which extends the forest of Elbeuf. A tribunal and chamber of commerce, a board of trade-arbitrators, a lycée, a branch of the Bank of France, a school of industry, a school of cloth manufacture and a museum of natural history are among its institutions. The churches of St Étienne and St Jean, both of the Renaissance period with later additions, preserve stained glass of the 16th century. The hôtel-de-ville and the Cercle du Commerce are the chief modern buildings. The town with its suburbs, Orival, Caudebec-lès-Elbeuf, St Aubin and St Pierre, is one of the principal and most ancient seats of the woollen manufacture in France; more than half the inhabitants are directly maintained by the staple industry and numbers more by the auxiliary crafts. As a river-port it has a brisk trade in the produce of the surrounding district as well as in the raw materials of its manufactures, especially in wool from La Plata, Australia and Germany. Two bridges, one of them a suspension-bridge, communicate with St Aubin on the opposite bank of the Seine, and steamboats ply regularly to Rouen.

Elbeuf was, in the 13th century, the centre of an important fief held by the house of Harcourt, but its previous history goes back at least to the early years of the Norman occupation, when it appears under the name of Hollebof. It passed into the hands of the houses of Rieux and Lorraine, and was raised to the rank of a duchy in the peerage of France by Henry III. in favour of Charles of Lorraine (d. 1605), grandson of Claude, duke of Guise, master of the hounds and master of the horse of France. The last duke of Elbeuf was Charles Eugène of Lorraine, prince de Lambesc, who distinguished himself in 1789 by his energy in repressing risings of the people at Paris. He fought in the army of the Bourbons, and later in the service of Austria, and died in 1825.

ELBING, a seaport town of Germany, in the kingdom of Prussia, 49 m. by rail E.S.E. of Danzig, on the Elbing, a small river which flows into the Frische Haff about 5 m. from the town, and is united with the Nogat or eastern arm of the Vistula by means of the Kraffohl canal. Pop. (1905) 55,627. By the Elbing-Oberländischer canal, 110 m. long, constructed in 1845-1860, Lakes Geserich and Drewenz are connected with Lake Drausen, and consequently with the port of Elbing. The old town was formerly surrounded by fortifications, but of these only a few fragments remain. There are several churches, among them the Marienkirche (dating from the 15th century and restored in 1887), a classical school (Gymnasium) founded in 1536, a modern school (Realschule), a public library of over 28,000 volumes, and several charitable institutions. The town-hall (1894) contains a historical museum.

Elbing is a place of rapidly growing industries. At the great Schichau iron-works, which employ thousands of workmen, are built most of the torpedo-boats and destroyers for the German navy, as well as larger craft, locomotives and machinery. In addition to this there are at Elbing important iron foundries, and manufactories of machinery, cigars, lacquer and metal ware, flax and hemp yarn, cotton, linen, organs, &c. There is a considerable trade also in agricultural produce.

The origin of Elbing was a colony of traders from Lübeck and Bremen, which established itself under the protection of a castle of the Teutonic Knights, built in 1237. In 1246 the town acquired “Lübeck rights,” *i.e.*
the full autonomy conceded by the charter of the emperor Frederick II. in 1226 (see Lübeck), and it was early 164
admitted to the Hanseatic League. In 1454 the town repudiated the overlordship of the Teutonic Order, and placed itself under the protection of the king of Poland, becoming the seat of a Polish voivode. From this event dates a decline in its prosperity, a decline hastened by the wars of the early 18th century. In 1698, and again in 1703, it was seized by the elector of Brandenburg as security for a debt due to him by the Polish king. It was taken and held to ransom by Charles XII. of Sweden, and in 1710 was captured by the Russians. In 1772, when it fell to Prussia through the first partition of Poland, it was utterly decayed.

See Fuchs, *Gesch. der Stadt Elbing* (Elbing, 1818-1852); Rhode, *Der Elbinger Kreis in topographischer, historischer, und statistischer Hinsicht* (Danzig, 1871); Wernick, *Elbing* (Elbing, 1888).

ELBOW, in anatomy, the articulation of the *humerus*, the bone of the upper arm, and the *ulna* and *radius*, the bones of the forearm (see Joints). The word is thus applied to things which are like this joint in shape, such as a sharp bend of a stream or river, an angle in a tube, &c. The word is derived from the O. Eng.

elnboga, a combination of *eln*, the forearm, and *boga*, a bow or bend. This combination is common to many Teutonic languages, cf. Ger. *Ellbogen*. *Eln* still survives in the name of a linear measure, the "ell," and is derived from the O. Teut. *alina*, cognate with Lat. *ulna* and Gr. ὠλένη, the forearm. The use of the arm as a measure of length is illustrated by the uses of *ulna*, in Latin, cubit, and fathom.

ELBURZ, or Alburz (from O. Pers. *Hara-bere-zaiti*, the "High Mountain"), a great chain of mountains in northern Persia, separating the Caspian depression from the Persian highlands, and extending without any break for 650 m. from the western shore of the Caspian Sea to north-eastern Khorasan. According to the direction, or strike, of its principal ranges the Elburz may be divided into three sections: the first 120 m. in length with a direction nearly N. to S., the second 240 m. in length with a direction N.W. to S.E., and the third 290 m. in length striking S.W. to N.E. The first section, which is connected with the system of the Caucasus, and begins west of Lenkoran in 39° N. and 45° E., is known as the Talish range and has several peaks 9000 to 10,000 ft. in height. It runs almost parallel to the western shore of the Caspian, and west of Astara is only 10 or 12 m. distant from the sea. At the point west of Resht, where the direction of the principal range changes to one of N.W. to S.E., the second section of the Elburz begins, and extends from there to beyond Mount Demavend, east of Teheran. South of Resht this section is broken through at almost a right angle by the Safid Rud (White river), and along it runs the principal commercial road between the Caspian and inner Persia, Resht-Kazvin-Teheran. The Elburz then splits into three principal ranges running parallel to one another and connected at many places by secondary ranges and spurs. Many peaks of the ranges in this section have an altitude of 11,000 to 13,000 ft., and the elevation of the passes leading over the ranges varies between 7000 and 10,000 ft. The highest peaks are situated in the still unexplored district of Talikan, N.W. of Teheran, and thence eastwards to beyond Mount Demavend. The part of the Elburz immediately north of Teheran is known as the Kuh i Shimran (mountain of Shimran, from the name of the Shimran district on its southern slopes) and culminates in the Sar i Tochal (12,600 ft.). Beyond it, and between the border of Talikan in the N.W. and Mount Demavend in the N.E., are the ranges Azadbur, Kasil, Kachang, Kendevan, Shahzad, Varzeh, Derbend i Sar and others, with elevations of 12,000 to 13,500 ft., while Demavend towers above them all with its altitude of 19,400 ft. The eastern foot of Demavend is washed by the river Herhaz (called Lar river in its upper course), which there breaks through the Elburz in a S.-N. direction in its course to the Caspian, past the city of Amol. The third section of the Elburz, with its principal ranges striking S.W. to N.E., has a length of about 290 m., and ends some distance beyond Bujnurd in northern Khorasan, where it joins the Ala Dagh range, which has a direction to the S.E., and, continuing with various appellations to northern Afghanistan, unites with the Paropamisus. For about two-thirds of its length—from its beginning to Khush Yailak—the third section consists of three principal ranges connected by lateral ranges and spurs. It also has many peaks over 10,000 ft. in height, and the Nizva mountain on the southern border of the unexplored district of Hazarjirib, north of Semnan, and the Shahkuh, between Shahrud and Astarabad, have an elevation exceeding 13,000 ft. Beyond Khush Yailak (meaning "pleasant summer quarters"), with an elevation of 10,000 ft., are the Kuh i Buhar (8000) and Kuh i Suluk (8000), which latter joins the Ala Dagh (11,000).

The northern slopes of the Elburz and the lowlands which lie between them and the Caspian, and together form the provinces of Gilan, Mazandaran and Astarabad, are covered with dense forest and traversed by hundreds (Persian writers say 1362) of perennial rivers and streams. The breadth of the lowlands between the foot of the hills and the sea is from 2 to 25 m., the greatest breadth being in the meridian of Resht in Gilan, and in the districts of Amol, Sari and Barfurush in Mazandaran. The inner slopes and ranges of the Elburz south of the principal watershed, generally the central one of the three principal ranges which are outside of the fertilizing influence of the moisture brought from the sea, have little or no natural vegetation, and those farthest south are, excepting a few stunted cypresses, completely arid and bare.

"North of the principal watershed forest trees and general verdure refresh the eye. Gurgling water, strips of sward and tall forest trees, backed by green hills, make a scene completely unlike the usual monotony of Persian landscape. The forest scenery much resembles that of England, with fine oaks and greensward. South of the watershed the whole aspect of the landscape is as hideous and disappointing as scenery in Afghanistan. Ridge after ridge of bare hill and curtain behind curtain of serrated mountain, certainly sometimes of charming greys and blues, but still all bare and naked, rugged and arid" ("Beresford Lovett, *Proc. R.G.S.*, Feb. 1883).

The higher ranges of the Elburz are snow-capped for the greater part of the year, and some, which are not exposed to the refracted heat from the arid districts of inner Persia, are rarely without snow. Water is plentiful in the Elburz, and situated in well-watered valleys and gorges are innumerable flourishing villages, embosomed in gardens and orchards, with extensive cultivated fields and meadows, and at higher altitudes small plateaus, under snow until March or April, afford cool camping grounds to the nomads of the plains, and luxuriant grazing to their sheep and cattle during the summer.

(A. H.-S.)

ELCHE, a town of eastern Spain, in the province of Alicante, on the river Vinalapo. Pop. (1900) 27,308. Elche is the meeting-place of three railways, from Novelda, Alicante and Murcia. It contains no building of high architectural merit, except, perhaps, the collegiate church of Santa Maria, with its lofty blue-tiled dome and fine west doorway. But the costume and physiognomy of the inhabitants, the narrow streets and flatroofed, whitewashed houses, and more than all, the thousands of palm-trees in its gardens and fields, give the place a strikingly Oriental aspect, and render it unique among the cities of Spain. The cultivation of the palm is indeed the principal occupation; and though the dates are inferior to those of the Barbary States, upwards of 22,500 tons are annually exported. The blanched fronds are also sold in large quantities for the processions of Palm Sunday, and after they have received the blessing of the priest they are regarded throughout Spain as certain defences against lightning. Other thriving local industries include the manufacture of oil, soap, flour, leather, alcohol and esparto grass rugs. The harbour of Elche is Santa Pola (pop. 4100), situated 6 m. E.S.E., where the Vinalapo enters the Mediterranean, after forming the wide lagoon known as the Albufera de Elche.

Elche is usually identified with the Iberian *Helike*, afterwards the Roman colony of *Ilici* or *Illici*. From the 8th century to the 13th it was held by the Moors, who finally failed to recapture it from the Spaniards in 1332.

ELCHINGEN, a village of Germany, in the kingdom of Bavaria, not far from the Danube, 5 m. N.E. from Ulm. Here, on the 14th of October 1805, the Austrians under Laudon were defeated by the French under 165
Ney, who by taking the bridge decided the day and gained for himself the title of duke of Elchingen.

ELDAD BEN MAḤLI, also surnamed had-Dani, Abu-Dani, David-had-Dani, or the Danite, Jewish traveller, was the supposed author of a Jewish travel-narrative of the 9th century A.D., which enjoyed great authority in the middle ages, especially on the question of the Lost Ten Tribes. Eldad first set out to visit his Hebrew brethren in Africa and Asia. His vessel was wrecked, and he fell into the hands of cannibals; but he was saved by his leanness, and by the opportune invasion of a neighbouring tribe. After spending four years with his new captors, he was ransomed by a fellow-countryman, a merchant of the tribe of Issachar. He then (according to his highly fabulous narrative) visited the territory of Issachar, in the mountains of Media and Persia; he also describes the abodes of Zabulon, on the "other side" of the Paran Mountains, extending to Armenia and the Euphrates; of Reuben, on another side of the same mountains; of Ephraim and Half Manasseh, in Arabia, not far from Mecca; and of Simeon and the other Half of Manasseh, in Chorazin, six months' journey from Jerusalem. Dan, he declares, sooner than join in Jeroboam's scheme of an Israelite war against Judah, had migrated to Cush, and finally, with the help of Naphthali, Asher and Gad, had founded an independent Jewish kingdom in the Gold Land of Havila, beyond Abyssinia. The tribe of Levi had also been miraculously guided, from near Babylon, to Havila, where they were enclosed and protected by the mystic river Sambation or Sabbation, which on the Sabbath, though calm, was veiled in impenetrable mist, while on other days it ran with a fierce untraversable current of stones and sand.

Apart from these tales, we have the genuine Eldad, a celebrated Jewish traveller and philologist; who flourished *c.* A.D. 830-890; to whom the work above noticed is ascribed; who was a native either of S. Arabia, Palestine or Media; who journeyed in Egypt, Mesopotamia, North Africa, and Spain; who spent several years at Kairawan in Tunis; who died on a visit to Cordova, and whose authority, as to the lost tribes, is supported by a great Hebrew doctor of his own time, Ẓemaḥ Gaon, the rector of the Academy at Sura (A.D. 889-898). It is possible that a certain relationship exists (as suggested by Epstein and supported by D.H. Müller) between the famous apocryphal *Letter of Prester John* (of *c.* A.D. 1165) and the narrative of Eldad; but the affinity is not close. Eldad is quoted as an authority on linguistic difficulties by the leading medieval Jewish grammarians and lexicographers.

The work ascribed to Eldad is in Hebrew, divided into six chapters, probably abbreviated from the original text. The first edition appeared at Mantua about 1480; the second at Constantinople in 1516; this was reprinted at Venice in 1544 and 1605, and at Jessnitz in 1722. A Latin version by Gilb. Génébrard was published at Paris in 1563, under the title of *Eldad Danius ... de Judaeis clausis eorumque in Aethiopia ... imperio*, and was afterwards incorporated in the translator's *Chronologia Hebraeorum* of 1584; a German version appeared at Prague in 1695, and another at Jessnitz in 1723. In 1838 E. Carmoly edited and translated a fuller recension which he had found in a MS. from the library of Eliezer Ben Hasan, forwarded to him by David Zabach of Morocco (see *Relation d'Eldad le Danite*, Paris, 1838). Both forms are printed by Dr Jellinek in his *Bet-ha-Midrasch*, vols. ii. p. 102, &c., and iii. p. 6, &c. (Leipzig, 1853-1855). See also Bartolocci, *Bibliotheca magna Rabbinica*, i. 101-130; Fürst, *Bibliotheca Judaica*, i. 30, &c.; Hirsch Graetz, *Geschichte der Juden* (3rd ed., Leipzig, 1895), v. 239-244; Rossi, *Dizionario degli Ebrei*; Steinschneider, *Cat. librorum Hebraeorum in bibliotheca Bodleiana*, cols. 923-925; Kitto's *Biblical Cyclopaedia* (3rd edition, *sub nomine*); Abr. Epstein, *Eldad ha-Dani* (Pressburg, 1891); D.H. Müller, "Die Recensionen und Versionen des Eldad had-Dani," in *Denkschriften d. Wiener Akad.* (Phil.-Hist. Cl.), vol. xli. (1892), pp. 1-80.

ELDER (Gr. πρεσβύτερος), the name given at different times to a ruler or officer in certain political and ecclesiastical systems of government.

1. The office of elder is in its origin political and is a relic of the old patriarchal system. The unit of primitive society is always the family; the only tie that binds men together is that of kinship. “The eldest male parent,” to quote Sir Henry Maine,[1] “is absolutely supreme in his household. His dominion extends to life and death and is as unqualified over his children and their houses as over his slaves.” The tribe, which is a later development, is always an aggregate of families or clans, not a collection of individuals. “The union of several clans for common political action,” as Robertson Smith says, “was produced by the pressure of practical necessity, and always tended towards dissolution when this practical pressure was withdrawn. The only organization for common action was that the leading men of the clans consulted together in time of need, and their influence led the masses with them. Out of these conferences arose the senates of elders found in the ancient states of Semitic and Aryan antiquity alike.”[2] With the development of civilization there came a time when age ceased to be an indispensable condition of leadership. The old title was, however, generally retained, *e.g.* the γέροντες so often mentioned in Homer, the γερουσία of the Dorian states, the *senatus* and the *patres conscripti* of Rome, the sheikh or elder of Arabia, the alderman of an English borough, the seigneur (Lat. *senior*) of feudal France.

2. It was through the influence of Judaism that the originally political office of elder passed over into the Christian Church and became ecclesiastical. The Israelites inherited the office from their Semitic ancestors (just as did the Moabites and the Midianites, of whose elders we read in Numbers xxii. 7), and traces of it are found throughout their history. Mention is made in Judges viii. 14 of the elders of Succoth whom “Gideon taught with thorns of the wilderness and with briers.” It was to the elders of Israel in Egypt that Moses communicated the plan of Yahweh for the redemption of the people (Exodus iii. 16). During the sojourn in the wilderness the elders were the intermediaries between Moses and the people, and it was out of the ranks of these elders that Moses chose a council of seventy “to bear with him the burden of the people” (Numbers xi. 16). The elders were the governors of the people and the administrators of justice. There are frequent references to their work in the latter capacity in the book of Deuteronomy, especially in relation to the following crimes—the disobedience of sons; slander against a wife; the refusal of levirate marriage; manslaughter; and blood-revenge. Their powers were gradually curtailed by (*a*) the development of the monarchy, to which of course they were in subjection, and which became the court of appeal in questions of law;[3] (*b*) the appointment of special judges, probably chosen from amongst the elders themselves, though their appointment meant the loss of privilege to the general body; (*c*) the rise of the priestly orders, which usurped many of the prerogatives that originally belonged to the elders. But in spite of the rise of new authorities, the elders still retained a large amount of influence. We hear of them frequently in the Persian, Greek and Roman periods. In the New Testament the members of the Sanhedrin in Jerusalem are very frequently termed “elders” or πρεσβύτεροι, and from them the name was taken over by the Church.

3. The name “elder” was probably the first title bestowed upon the officers of the Christian Church—since the word deacon does not occur in connexion with the appointment of the Seven in Acts vi. Its universal adoption is due not only to its currency amongst the Jews, but also to the fact that it was frequently used as the title of magistrates in the cities and villages of Asia Minor. For the history of the office of elder in the early Church and the relation between elders and bishops see Presbyter.

4. In modern times the use of the term is almost entirely confined to the Presbyterian church, the officers of which are always called elders. According to the Presbyterian theory of church government there are two classes of elders—“teaching elders,” or those specially set apart to the pastoral office, and “ruling elders,” who are laymen, chosen generally by the congregation and set apart by ordination to be associated with the pastor in the oversight and government of the church. When the word is used without any qualification it is 166
understood to apply to the latter class alone. For an account of the duties, qualifications and powers of elders in the Presbyterian Church see Presbyterianism.

See W.R. Smith, *History of the Semites*; H. Maine, *Ancient Law*; E. Schürer, *The Jewish People in the Time of Christ*; J. Wellhausen, *History of Israel and Judah*; G.A. Deissmann, *Bible Studies*, p. 154.

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1. *Ancient Law*, p. 126.

2. *Religion of the Semites*, p. 34.

3. There is a hint at this even in the Pentateuch, “every great matter they shall bring unto thee, but every small matter they shall judge themselves.”

ELDER (O. Eng. *ellarn*; Ger. *Holunder*; Fr. *sureau*), the popular designation of the deciduous shrubs and trees constituting the genus *Sambucus* of the natural order Caprifoliaceae. The Common Elder, *S. nigra*, the bourtree of Scotland, is found in Europe, the north of Africa, Western Asia, the Caucasus, and Southern Siberia; in sheltered spots it attains a height of over 20 ft. The bark is smooth; the shoots are stout and angular, and the leaves glabrous, pinnate, with oval or elliptical leaflets. The flowers, which form dense flat-topped clusters (corymbose cymes), with five main branches, have a cream-coloured, gamopetalous, five-lobed corolla, five stamens, and three sessile stigmas; the berries are purplish-black, globular and three- or four-seeded, and ripen about September. The elder thrives best in moist, well-drained situations, but can be grown in a great diversity of soils. It grows readily from young shoots, which after a year are fit for transplantation. It is found useful for making screen-fences in bleak, exposed situations, and also as a shelter for other shrubs in the outskirts of plantations. By clipping two or three times a year, it may be made close and compact in growth. The young trees furnish a brittle wood, containing much pith; the wood of old trees is white, hard and close-grained, polishes well, and is employed for shoemakers’ pegs, combs, skewers, mathematical instruments and turned articles. Young elder twigs deprived of pith have from very early times been in request for making whistles, popguns and other toys.

The elder was known to the ancients for its medicinal properties, and in England the inner bark was formerly administered as a cathartic. The flowers (*sambuci flores*) contain a volatile oil, and serve for the distillation of elder-flower water (*aqua sambuci*), used in confectionery, perfumes and lotions. The leaves of the elder are employed to impart a green colour to fat and oil (*unguentum sambuci foliorum* and *oleum viride*), and the berries for making wine, a common adulterant of port. The leaves and bark emit a sickly odour, believed to be repugnant to insects. Christopher Gullet (*Phil. Trans.*, 1772, lxii. p. 348) recommends that cabbages, turnips, wheat and fruit trees, to preserve them from caterpillars, flies and blight, should be whipped with twigs of young elder. According to German folklore, the hat must be doffed in the presence of the elder-tree; and in certain of the English midland counties a belief was once prevalent that the cross of Christ was made from its wood, which should therefore never be used as fuel, or treated with disrespect (see *Quart. Rev.* cxiv. 233). It was, however, a common medieval tradition, alluded to by Ben Jonson, Shakespeare and other writers, that the elder was the tree on which Judas hanged himself; and on this account, probably, to be crowned with elder was in olden times accounted a disgrace. In Cymbeline (act iv. s. 2) “the stinking elder” is mentioned as a symbol of grief. In Denmark the tree is supposed by the superstitious to be under the protection of the “Elder-mother”: its flowers may not be gathered without her leave; its wood must not be employed for any household furniture; and a child sleeping in an elder-wood cradle would certainly be strangled by the Elder-mother.

Several varieties are known in cultivation: *aurea*, golden elder, has golden-yellow leaves; *laciniata*, parsley-leaved elder, has the leaflets cut into fine segments; *rotundifolia* has rounded leaflets; forms also occur with variegated white and yellow leaves, and *virescens* is a variety having white bark and green-coloured berries. The scarlet-berried elder, *S. racemosa*, is the handsomest species of the genus. It is a native of various parts of Europe, growing in Britain to a height of over 15 ft., but often producing no fruit. The dwarf elder or Danewort (supposed to have been introduced into Britain by the Danes), *S. Ebulus*, a common European species, reaches a height of about 6 ft. Its cyme is hairy, has three principal branches, and is smaller than that of *S. nigra*; the flowers are white tipped with pink. All parts of the plant are cathartic and emetic.

ELDON, JOHN SCOTT, 1st Earl of (1751-1838), lord high chancellor of England, was born at Newcastle on the 4th of June 1751. His grandfather, William Scott of Sandgate, a suburb of Newcastle, was clerk to a “fitter”—a sort of water-carrier and broker of coals. His father, whose name also was William, began life as an apprentice to a fitter, in which service he obtained the freedom of Newcastle, becoming a member of the gild of Hoastmen (coal-fitters); later in life he became a principal in the business, and attained a respectable position as a merchant in Newcastle, accumulating property worth nearly £20,000.

John Scott was educated at the grammar school of his native town. He was not remarkable at school for application to his studies, though his wonderful memory enabled him to make good progress in them; he frequently played truant and was whipped for it, robbed orchards, and indulged in other questionable schoolboy freaks; nor did he always come out of his scrapes with honour and a character for truthfulness. When he had finished his education at the grammar school, his father thought of apprenticing him to his own business, to which an elder brother Henry had already devoted himself; and it was only through the interference of his elder brother William (afterwards Lord Stowell, *q.v.*), who had already obtained a fellowship at University College, Oxford, that it was ultimately resolved that he should continue the prosecution of his studies. Accordingly, in 1766, John Scott entered University College with the view of taking holy orders and obtaining a college living. In the year following he obtained a fellowship, graduated B.A. in 1770, and in 1771 won the prize for the English essay, the only university prize open in his time for general competition.

His wife was the eldest daughter of Aubone Surtees, a Newcastle banker. The Surtees family objected to the match, and attempted to prevent it; but a strong attachment had sprung up between them. On the 18th November 1772 Scott, with the aid of a ladder and an old friend, carried off the lady from her father’s house in the Sandhill, across the border to Blackshiels, in Scotland, where they were married. The father of the bridegroom objected not to his son’s choice, but to the time he chose to marry; for it was a blight on his son’s prospects, depriving him of his fellowship and his chance of church preferment. But while the bride’s family refused to hold intercourse with the pair, Mr Scott, like a prudent man and an affectionate father, set himself to make the best of a bad matter, and received them kindly, settling on his son £2000. John returned with his wife to Oxford, and continued to hold his fellowship for what is called the year of grace given after marriage, and added to his income by acting as a private tutor. After a time Mr Surtees was reconciled with his daughter, and made a liberal settlement on her.

John Scott’s year of grace closed without any college living falling vacant; and with his fellowship he gave up the church and turned to the study of law. He became a student at the Middle Temple in January 1773. In 1776 he was called to the bar, intending at first to establish himself as an advocate in his native town, a scheme which his early success led him to abandon, and he soon settled to the practice of his profession in London, and on the northern circuit. In the autumn of the year in which he was called to the bar his father died, leaving him a legacy of £1000 over and above the £2000 previously settled on him.

In his second year at the bar his prospects began to brighten. His brother William, who by this time held the Camden professorship of ancient history, and enjoyed an extensive acquaintance with men of eminence in London, was in a position materially to advance his interests. Among his friends was the notorious Andrew Bowes of Gibside, to the patronage of whose house the rise of the Scott family was largely owing. Bowes having contested Newcastle and lost it, presented an election petition against the return of his opponent. Young Scott was retained as junior counsel in the case, and though he lost the petition he did not fail to improve the opportunity which it afforded for displaying his talents. This engagement, in the commencement of his second year at the bar, and the dropping in of occasional fees, must have raised his hopes; and he now abandoned the scheme of becoming a provincial barrister. A year or two of dull drudgery and few fees followed, and he began to be much depressed. But in 1780 we find his prospects suddenly improved, by his appearance in the case of *Ackroyd* v. *Smithson*, which became a leading case settling a rule of law; and young Scott, having lost his point in the inferior court, insisted on arguing it, on appeal, against the opinion of his clients, and carried it before Lord Thurlow, whose favourable consideration he won by his able argument. The same year Bowes again retained him in an election petition; and in the year following Scott greatly increased his reputation by his appearance as leading counsel in the Clitheroe election petition. From this time his success was certain. In 1782 he obtained a silk gown, and was so far cured of his early modesty that he declined accepting the king's counselship if precedence over him were given to his junior, Thomas (afterwards Lord) Erskine, though the latter was the son of a peer and a most accomplished orator. He was now on the high way to fortune. His health, which had hitherto been but indifferent, strengthened with the demands made upon it; his talents, his power of endurance, and his ambition all expanded together. He enjoyed a considerable practice in the northern part of his circuit, before parliamentary committees and at the chancery bar. By 1787 his practice at the equity bar had so far increased that he was obliged to give up the eastern half of his circuit (which embraced six counties) and attend it only at Lancaster.

In 1782 he entered parliament for Lord Weymouth's close borough of Weobley, which Lord Thurlow obtained for him without solicitation. In parliament he gave a general and independent support to Pitt. His first parliamentary speeches were directed against Fox's India Bill. They were unsuccessful. In one he aimed at being brilliant; and becoming merely laboured and pedantic, he was covered with ridicule by Sheridan, from whom he received a lesson which he did not fail to turn to account. In 1788 he was appointed solicitor-general, and was knighted, and at the close of this year he attracted attention by his speeches in support of Pitt's resolutions on the state of the king (George III., who then laboured under a mental malady) and the delegation of his authority. It is said that he drew the Regency Bill, which was introduced in 1789. In 1793 Sir John Scott was promoted to the office of attorney-general, in which it fell to him to conduct the memorable prosecutions for high treason against British sympathizers with French republicanism,—amongst others, against the celebrated Horne Tooke. These prosecutions, in most cases, were no doubt instigated by Sir John Scott, and were the most important proceedings in which he was ever professionally engaged. He has left on record, in his *Anecdote Book*, a defence of his conduct in regard to them. A full account of the principal trials, and of the various legislative measures for repressing the expressions of popular opinion for which he was more or less responsible, will be found in Twiss's *Public and Private Life of the Lord Chancellor Eldon*, and in the *Lives of the Lord Chancellors*, by Lord Campbell.

In 1799 the office of chief justice of the Court of Common Pleas falling vacant, Sir John Scott's claim to it was not overlooked; and after seventeen years' service in the Lower House, he entered the House of Peers as Baron Eldon. In February 1801 the ministry of Pitt was succeeded by that of Addington, and the chief justice now ascended the woolsack. The chancellorship was given to him professedly on account of his notorious anti-Catholic zeal. From the peace of Amiens (1802) till 1804 Lord Eldon appears to have interfered little in politics. In the latter year we find him conducting the negotiations which resulted in the dismissal of Addington and the recall of Pitt to office as prime minister. Lord Eldon was continued in office as chancellor under Pitt; but the new administration was of short duration, for on the 23rd of January 1806 Pitt died, worn out with the anxieties of office, and his ministry was succeeded by a coalition, under Lord Grenville. The death of Fox, who became foreign secretary and leader of the House of Commons, soon, however, broke up the Grenville administration; and in the spring of 1807 Lord Eldon once more, under Lord Liverpool's administration, returned to the woolsack, which, from that time, he continued to occupy for about twenty years, swaying the cabinet, and being in all but name prime minister of England. It was not till April 1827, when the premiership, vacant through the paralysis of Lord Liverpool, fell to Canning, the chief advocate of Roman Catholic emancipation, that Lord Eldon, in the seventy-sixth year of his age, finally resigned the chancellorship. When, after the two short administrations of Canning and Goderich, it fell to the duke of Wellington to construct a cabinet, Lord Eldon expected to be included, if not as chancellor, at least in some important office, but he was overlooked, at which he was much chagrined. Notwithstanding his frequent protests that he did not covet power, but longed for retirement, we find him again, so late as 1835, within three years of his death, in hopes of office under Peel. He spoke in parliament for the last time in July 1834.

In 1821 Lord Eldon had been created Viscount Encombe and earl of Eldon by George IV., whom he managed to conciliate, partly, no doubt, by espousing his cause against his wife, whose advocate he had formerly been, and partly through his reputation for zeal against the Roman Catholics. In the same year his brother William, who from 1798 had filled the office of judge of the High Court of Admiralty, was raised to the peerage under the title of Lord Stowell.

Lord Eldon's wife, his dear "Bessy," his love for whom is a beautiful feature in his life, died before him, on the 28th of June 1831. By nature she was of simple character, and by habits acquired during the early portion of her husband's career almost a recluse. Two of their sons reached maturity—John, who died in 1805, and William Henry John, who died unmarried in 1832. Lord Eldon himself survived almost all his immediate relations. His brother William died in 1836. He himself died in London on the 13th of January 1838, leaving behind him two daughters, Lady Frances Bankes and Lady Elizabeth Repton, and a grandson John (1805-1854), who succeeded him as second earl, the title subsequently passing to the latter's son John (b. 1846).

Lord Eldon was no legislator—his one aim in politics was to keep in office, and maintain things as he found them; and almost the only laws he helped to pass were laws for popular coercion. For nearly forty years he fought against every improvement in law, or in the constitution—calling God to witness, on the smallest proposal of reform, that he foresaw from it the downfall of his country. Without any political principles, properly so called, and without interest in or knowledge of foreign affairs, he maintained himself and his party in power for an unprecedented period by his great tact, and in virtue of his two great political properties—of zeal against every species of reform, and zeal against the Roman Catholics. To pass from his political to his judicial character is to shift to ground on which his greatness is universally acknowledged. His judgments, which have received as much praise for their accuracy as abuse for their clumsiness and uncouthness, fill a small library. But though intimately acquainted with every nook and cranny of the English law, he never carried his studies into foreign fields, from which to enrich our legal literature; and it must be added that against the excellence of his judgments, in too many cases, must be set off the hardships, worse than injustice, that arose from his protracted delays in pronouncing them. A consummate judge and the narrowest of politicians, he was doubt on the bench, and promptness itself in the political arena. For literature, as for art, he had no feeling. What intervals of leisure he enjoyed from the cares of office he filled up with newspapers and the gossip of old cronies. Nor were his intimate associates men of refinement and taste; they were rather good fellows who quietly enjoyed a good bottle and a joke; he uniformly avoided encounters of wit with his equals. He is said to have been parsimonious, and certainly he was quicker to receive than to reciprocate hospitalities; but his mean establishment and mode of life are explained by the retired habits of his wife, and her dislike of company. His manners were very winning and courtly, and in the 168 circle of his immediate relatives he is said to have always been lovable and beloved.

"In his person," says Lord Campbell, "Lord Eldon was about the middle size, his figure light and athletic, his features regular and handsome, his eye bright and full, his smile remarkably benevolent, and his whole appearance prepossessing. The advance of years rather increased than detracted from these personal advantages. As he sat on the judgment-seat, 'the deep thought betrayed in his furrowed brow—the large eyebrows, overhanging eyes that seemed to regard more what was taking place within than around him—his calmness, that would have assumed a character of sternness but for its perfect placidity—his dignity, repose and venerable age, tended at once to win confidence and to inspire respect' (Townsend). He had a voice both sweet and deep-toned, and its effect was not injured by his Northumbrian burr, which, though strong, was entirely free from harshness and vulgarity."

Authorities.—Horace Twiss, *Life of Lord Chancellor Eldon* (1844); W.E. Surtees, *Sketch of the Lives of Lords Stowell and Eldon* (1846); Lord Campbell, *Lives of the Chancellors*; W.C. Townsend, *Lives of Twelve Eminent Judges* (1846); *Greville Memoirs*.

EL DORADO (Span. "the gilded one"), a name applied, first, to the king or chief priest of a South American tribe who was said to cover himself with gold dust at a yearly religious festival held near Santa Fé de Bogotá; next, to a legendary city called Manoa or Omoa; and lastly, to a mythical country in which gold and precious stones were found in fabulous abundance. The legend, which has never been traced to its ultimate source, had many variants, especially as regards the situation attributed to Manoa. It induced many Spanish explorers to lead expeditions in search of treasure, but all failed. Among the most famous were the expedition undertaken by Diego de Ordaz, whose lieutenant Martinez claimed to have been rescued from shipwreck, conveyed inland, and entertained at Omoa by "El Dorado" himself (1531); and the journeys of Orellana (1540-1541), who passed down the Rio Napo to the valley of the Amazon; that of Philip von Hutten (1541-1545), who led an exploring party from Coro on the coast of Caracas; and of Gonzalo Ximenes de Quesada (1569), who started from Santa Fé de Bogotá. Sir Walter Raleigh, who resumed the search in 1595, described Manoa as a city on Lake Parimá in Guiana. This lake was marked on English and other maps until its existence was disproved by A. von Humboldt (1769-1859). Meanwhile the name of El Dorado came to be used metaphorically of any place where wealth could be rapidly acquired. It was given to a county in California, and to towns and cities in various states. In literature frequent allusion is made to the legend, perhaps the best-known references being those in Milton's *Paradise Lost* (vi. 411) and Voltaire's *Candide* (chs. 18, 19).

See A.F.A. Bandelier, *The Gilded Man, El Dorado* (New York, 1893).

ELDUAYEN, JOSÉ DE, 1st Marquis del Pazo de la Merced (1823-1898), Spanish politician, was born in Madrid on the 22nd of June 1823. He was educated in the capital, took the degree of civil engineer, and as such directed important works in Asturias and Galicia, entered the Cortes in 1856 as deputy for Vigo, and sat in all the parliaments until 1867 as member of the Union Liberal with Marshal O'Donnell. He attacked the Miraflores cabinet in 1864, and became under-secretary of the home office when Canovas was minister in 1865. He was made a councillor of state in 1866, and in 1868 assisted the other members of the Union Liberal in preparing the revolution. In the Cortes of 1872 he took much part in financial debates. He accepted office as member of the last Sagasta cabinet under King Amadeus. On the proclamation of the republic Elduayen very earnestly co-operated in the Alphonsist conspiracy, and endeavoured to induce the military and politicians to work together. He went abroad to meet and accompany the prince after the *pronunciamiento* of Marshal Campos, landed with him at Valencia, was made governor of Madrid, a marquis, grand cross of Charles III., and minister for the colonies in 1878. He accepted the portfolio of foreign affairs in the Canovas cabinet from 1883 to 1885, and was made a life senator. He always prided himself on having been one of the five members of the Cortes of 1870 who voted for Alphonso XII. when that parliament elected Amadeus of Savoy. He died at Madrid on the 24th of June 1898.

ELEANOR OF AQUITAINE (*c.* 1122-1204), wife of the English king Henry II., was the daughter and heiress of Duke William X. of Aquitaine, whom she succeeded in April 1137. In accordance with arrangements made by her father, she at once married Prince Louis, the heir to the French crown, and a month later her husband became king of France under the title of Louis VII. Eleanor bore Louis two daughters but no sons. This was probably the reason why their marriage was annulled by mutual consent in 1151, but contemporary scandal-mongers attributed the separation to the king's jealousy. It was alleged that, while accompanying her husband on the Second Crusade (1146-1149), Eleanor had been unduly familiar with her uncle, Raymond of Antioch. Chronology is against this hypothesis, since Louis and she lived on good terms together for two years after the Crusade. There is still less ground for the supposition that Henry of Anjou, whom she married immediately after the divorce, had been her lover before it. This second marriage, with a youth some years her junior, was purely political. The duchy of Aquitaine required a strong ruler, and the union with Anjou was eminently desirable. Louis, who had hoped that Aquitaine would descend to his daughters, was mortified and alarmed by the Angevin marriage; all the more so when Henry of Anjou succeeded to the English crown in 1154. From this event dates the beginning of the secular strife between England and France which runs like a red thread through medieval history.

Eleanor bore to her second husband five sons and three daughters; John, the youngest of their children, was born in 1167. But her relations with Henry passed gradually through indifference to hatred. Henry was an unfaithful husband, and Eleanor supported her sons in their great rebellion of 1173. Throughout the latter years of the reign she was kept in a sort of honourable confinement. It was during her captivity that Henry formed his connexion with Rosamond Clifford, the Fair Rosamond of romance. Eleanor, therefore, can hardly have been responsible for the death of this rival, and the romance of the poisoned bowl appears to be an invention of the next century.

Under the rule of Richard and John the queen became a political personage of the highest importance. To both her sons the popularity which she enjoyed in Aquitaine was most valuable. But in other directions also she did good service. She helped to frustrate the conspiracy with France which John concocted during Richard's captivity. She afterwards reconciled the king and the prince, thus saving for John the succession which he had forfeited by his misconduct. In 1199 she crushed an Angevin rising in favour of John's nephew, Arthur of Brittany. In 1201 she negotiated a marriage between her grand-daughter, Blanche of Castile, and Louis of France, the grandson of her first husband. It was through her staunch defence of Mirabeau in Poitou that John got possession of his nephew's person. She died on the 1st of April 1204, and was buried at Fontevrault. Although a woman of strong passions and great abilities she is, historically, less important as an individual than as the heiress of Aquitaine, a part of which was, through her second marriage, united to England for some four hundred years.

See the chronicles cited for the reigns of Henry II., Richard I. and John. Also Sir J.H. Ramsay, *Angevin Empire* (London, 1903); K. Norgate, *England under the Angevin Kings* (London, 1887); and A. Strickland, *Lives of the Queens of England*, vol. i. (1841).

(H. W. C. D.)

ELEATIC SCHOOL, a Greek school of philosophy which came into existence towards the end of the 6th century B.C., and ended with Melissus of Samos (fl. *c.* 450 B.C.). It took its name from Elea, a Greek city of lower Italy, the home of its chief exponents, Parmenides and Zeno. Its foundation is often attributed to Xenophanes of Colophon, but, although there is much in his speculations which formed part of the later Eleatic doctrine, it is probably more correct to regard Parmenides as the founder of the school. At all events, it was Parmenides who gave it its fullest development. The main doctrines of the Eleatics were evolved in opposition, on the one hand, to the physical theories of the early physical philosophers who explained all existence in terms of primary matter (see IONIAN SCHOOL), and, on the other hand, to the theory of Heraclitus that all existence may be summed up as perpetual change. As against these theories the Eleatics maintained that the true explanation of things lies in the conception of a universal unity of being. The senses with their changing and inconsistent reports cannot cognize this unity; it is by thought alone that we can pass beyond the false appearances of sense and arrive at the knowledge of being, at the fundamental truth that "the All is One." There can be no creation, for being cannot come from not-being; a thing cannot arise from that which is different from it. The errors of common opinion arise to a great extent from the ambiguous use of the verb "to be," which may imply existence or be merely the copula which connects subject and predicate. 169

In these main contentions the Eleatic school achieved a real advance, and paved the way to the modern conception of metaphysics. Xenophanes in the middle of the 6th century had made the first great attack on the crude mythology of early Greece, including in his onslaught the whole anthropomorphic system enshrined in the poems of Homer and Hesiod. In the hands of Parmenides this spirit of free thought developed on metaphysical lines. Subsequently, whether from the fact that such bold speculations were obnoxious to the general sense of propriety in Elea, or from the inferiority of its leaders, the school degenerated into verbal disputes as to the possibility of motion, and similar academic trifling. The best work of the school was absorbed in the Platonic metaphysic (see E. Caird, *Evolution of Theology in the Greek Philosophers*, 1904).

See further the articles on XENOPHANES; PARMENIDES; ZENO (of Elea); MELISSUS, with the works there quoted; also the histories of philosophy by Zeller, Gomperz, Windelband, &c.

ELECAMPANE (Med. Lat. *Enula Campana*), a perennial composite plant, the *Inula Helenium* of botanists, which is common in many parts of Britain, and ranges throughout central and southern Europe, and in Asia as far eastwards as the Himalayas. It is a rather rigid herb, the stem of which attains a height of from 3 to 5 ft.; the leaves are large and toothed, the lower ones stalked, the rest embracing the stem; the flowers are yellow, 2 in. broad, and have many rays, each three-notched at the extremity. The root is thick, branching and mucilaginous, and has a warm, bitter taste and a camphoraceous odour. For medicinal purposes it should be procured from plants not more than two or three years old. Besides *inulin*, $C_{12}H_{20}O_{10}$, a body isomeric with starch, the root contains *helenin*, C_6H_8O, a stearoptene, which may be prepared in white acicular crystals, insoluble in water, but freely soluble in alcohol. When freed from the accompanying inula-camphor by repeated crystallization from alcohol, helenin melts at 110° C. By the ancients the root was employed both as a medicine and as a condiment, and in England it was formerly in great repute as an aromatic tonic and stimulant of the secretory organs. "The fresh roots of elecampane preserved with sugar, or made into a syrup or conserve," are recommended by John Parkinson in his *Theatrum Botanicum* as "very effectual to warm a cold and windy stomack, and the pricking and stitches therein or in the sides caused by the Spleene, and to helpe the cough, shortnesse of breath, and wheesing in the Lungs." As a drug, however, the root is now seldom resorted to except in veterinary practice, though it is undoubtedly possessed of antiseptic properties. In France and Switzerland it is used in the manufacture of absinthe.

ELECTION (from Lat. *eligere*, to pick out), the method by which a choice or selection is made by a constituent body (the electors or electorate) of some person to fill a certain office or dignity. The procedure itself is called an election. Election, as a special form of selection, is naturally a loose term covering many subjects; but except in the theological sense (the doctrine of election), as employed by Calvin and others, for the choice by God of His "elect," the legal sense (see ELECTION, *in law*, below), and occasionally as a synonym for personal choice (one's own "election"), it is confined to the selection by the preponderating vote of some properly constituted body of electors of one of two or more candidates, sometimes for admission only to some private social position (as in a club), but more particularly in connexion with public representative positions in political government. It is thus distinguished from arbitrary methods of appointment, either where the right of nominating rests in an individual, or where pure chance (such as selection by lot) dictates the result. The part played by different forms of election in history is alluded to in numerous articles in this work, dealing with various countries and various subjects. It is only necessary here to consider certain important features in the elections, as ordinarily understood, namely, the exercise of the right of voting for political and municipal offices in the United Kingdom and America. See also the articles PARLIAMENT; REPRESENTATION; VOTING; BALLOT, &c., and UNITED STATES: *Political Institutions*. For practical details as to the conduct of political elections in England reference must be made to the various text-books on the subject; the candidate and his election agent require to be on their guard against any false step which might invalidate his return.

Law in the United Kingdom.—Considerable alterations have been made in recent years in the law of Great Britain and Ireland relating to the procedure at parliamentary and municipal elections, and to election petitions.

As regards parliamentary elections (which may be either the "general election," after a dissolution of parliament, or "by-elections," when casual vacancies occur during its continuance), the most important of the amending statutes is the Corrupt and Illegal Practices Act 1883. This act, and the Parliamentary Elections Act 1868, as amended by it, and other enactments dealing with corrupt practices, are temporary acts requiring annual renewal. As regards municipal elections, the Corrupt Practices (Municipal Elections) Act 1872 has been repealed by the Municipal Corporations Act 1882 for England, and by the Local Government (Ireland) Act 1898 for Ireland. The governing enactments for England are now the Municipal Corporations Act 1882, part iv., and the Municipal Elections (Corrupt and Illegal Practices) Act 1884, the latter annually renewable. The provisions of these enactments have been applied with necessary modifications to municipal and other local government elections in Ireland by orders of the Irish Local Government Board made under powers conferred by the Local Government (Ireland) Act 1898. In Scotland the law regulating municipal and other local government elections is now to be found in the Elections (Scotland) (Corrupt and Illegal Practices) Act 1890.

The alterations in the law have been in the direction of greater strictness in regard to the conduct of elections, and increased control in the public interest over the proceedings on election petitions. Various acts and payments which were previously lawful in the absence of any corrupt bargain or motive are now altogether forbidden under the name of "illegal practices" as distinguished from "corrupt practices." Failure on the part of a parliamentary candidate or his election agent to comply with the requirements of the law in any particular is sufficient to invalidate the return (see the articles BRIBERY and CORRUPT PRACTICES). Certain relaxations are, however, allowed in consideration of the difficulty of absolutely avoiding all deviation from the strict rules laid down. Thus, where the judges who try an election petition report that there has been treating, undue influence, or any illegal practice by the candidate or his election agent, but that it was trivial, unimportant and of a limited character, and contrary to the orders and without the sanction or connivance of the candidate or his election agent, and that the candidate and his election agent took all reasonable means for preventing corrupt and illegal practices, and that the election was otherwise free from such practices on their part, the election will not be avoided. The court has also the power to relieve from the consequences of certain innocent contraventions of the law caused by inadvertence or miscalculation.

The inquiry into a disputed parliamentary election was formerly conducted before a committee of the House of Commons, chosen as nearly as possible from both sides of the House for that particular business. The decisions of these tribunals laboured under the suspicion of being prompted by party feeling, and by an act of 1868 the jurisdiction was finally transferred to judges of the High Court, notwithstanding the general unwillingness of the bench to accept a class of business which they feared might bring their integrity into dispute. Section 11 of the act ordered, *inter alia*, that the trial of every election petition shall be conducted before a *puisne judge* of one of the common law courts at Westminster and Dublin; that the said courts shall each select a judge to be placed on the rota for the trial of election petitions; that the said judges shall try petitions standing for trial according to seniority or otherwise, as they may agree; that the trial shall take place in the county or borough to which the petition refers, unless the court should think it desirable to hold it elsewhere. The judge shall determine "whether the member whose return is complained of, or any and what other person, was duly returned and elected, or whether the election was void," and shall certify his determination to the speaker. When corrupt practices have been charged the judge shall also report (1) whether any such practice has been committed by or with the knowledge or consent of any candidate, and the nature thereof; (2) the names of persons proved to have been guilty of any corrupt practice; and (3) whether corrupt practices have extensively prevailed at the election. Questions of law were to be referred to the decision of the court of common pleas. On the abolition of that court by the Judicature Act 1873, the jurisdiction was transferred to the common pleas division, and again on the abolition of that division was transferred to the king's bench division, in whom it is now vested. The rota of judges for the trial of election petitions is also supplied by the king's bench division. The trial now takes place before two judges instead of one; and, when necessary, the number of judges on the rota may be increased. Both the judges who try a petition are to sign the certificates to be made to the speaker. If they differ as to the validity of a return, they are to state such difference in their certificate, and the return is to be held good; if they differ as to a report on any other matter, they are to certify their difference and make no report on such matter. The director of public prosecutions attends the trial personally or by representative. It is his duty to watch the proceedings in the public interest, to issue summonses to witnesses whose evidence is desired by the court, and to prosecute before the election court or elsewhere those persons whom he thinks to have been guilty of corrupt or illegal practices at the election in question. If an application is made for leave to withdraw a petition, copies of the affidavits in support are to be delivered to him; and he is entitled to be heard and to call evidence in opposition to such application. Witnesses are not excused from answering criminating questions; but their evidence cannot be used against them in any proceedings except criminal proceedings for perjury in respect of that evidence. If a witness answers truly all questions which he is required by the court to answer, he is entitled to receive a certificate of indemnity, which will save him from all proceedings for any offence under the Corrupt Practices Acts committed by him before the date of the certificate at or in relation to the election, except proceedings to enforce any incapacity incurred by such offence. An application for leave to withdraw a petition must be supported by affidavits from all the parties to the petition and their solicitors, and by the election agents of all of the parties who were candidates at the election. Each of these affidavits is to state that to the best of the deponent's knowledge and belief there has been no agreement and no terms or undertaking made or entered into as to the withdrawal, or, if any agreement has been made, shall state its terms. The applicant and his solicitor are also to state in their affidavits the grounds on which the petition is sought to be withdrawn. If any person makes an agreement for the withdrawal of a petition in consideration of a money payment, or of the promise that the seat shall be vacated or another petition withdrawn, or omits to state in his affidavit that he has made an agreement, lawful or unlawful, for the withdrawal, he is guilty of an indictable misdemeanour. The report of the judges to the speaker is to contain particulars as to illegal practices similar to those previously required as to corrupt practices; and they are to report further whether any candidate has been guilty by his agents of an illegal practice, and whether certificates of indemnity have been given to persons reported guilty of corrupt or illegal practices.

Election petitions.

The Corrupt Practices Acts apply, with necessary variations in details, to parliamentary elections in Scotland and Ireland.

The amendments in the law as to municipal elections are generally similar to those which have been made in parliamentary election law. The procedure on trial of petitions is substantially the same, and wherever no other provision is made by the acts or rules the procedure on the trial of parliamentary election petitions is to be followed. Petitions against municipal elections were dealt with in 35 & 36 Vict. c. 60. The election judges appoint a number of barristers, not exceeding five, as commissioners to try such petitions. No barrister can be appointed who is of less than fifteen years' standing, or a member of parliament, or holder of any office of profit (other than that of recorder) under the crown; nor can any barrister try a petition in any borough in which he is recorder or in which he resides, or which is included in his circuit. The barrister sits without a jury. The provisions are generally similar to those relating to parliamentary elections. The petition may allege that the election was avoided as to the borough or ward on the ground of general bribery, &c., or that the election of the person petitioned against was avoided by corrupt practices, or by personal disqualification, or that he had not the majority of lawful votes. The commissioner who tries a petition sends to the High Court a certificate of the result, together with reports as to corrupt and illegal practices, &c., similar to those made to the speaker by the judges who try a parliamentary election petition. The Municipal Elections (Corrupt and Illegal Practices) Act 1884 applied to school board elections subject to certain variations, and has been extended by the Local Government Act 1888 to county council elections, and by the Local Government Act 1894 to elections by parochial electors. The law in Scotland is on the same lines, and extends to all non-parliamentary elections, and, as has been stated, the English statutes have been applied with adaptations to all municipal and local government elections in Ireland.

United States.—Elections are much more frequent in the United States than they are in Great Britain, and they are also more complicated. The terms of elective officers are shorter; and as there are also more offices to be filled, the number of persons to be voted for is necessarily much greater. In the year of a presidential election the citizen may be called upon to vote at one time for all of the following: (1) National candidates—president and vice-president (indirectly through the electoral college) and members of the House of Representatives; (2) state candidates—governor, members of the state legislature, attorney-general, treasurer, &c.; (3) county candidates—sheriff, county judges, district attorney, &c.; (4) municipal or town candidates—mayor, aldermen, selectmen, &c. The number of persons actually voted for may therefore be ten or a dozen, or it may be many more. In addition, the citizen is often called upon to vote yea or nay on questions such as amendments to the state constitutions, granting of licences, and approval or disapproval of new municipal undertakings. As there may be, and generally is, more than one candidate for each office, and as all elections are now, and have been for many years, conducted by ballot, the total number of names to appear on the ballot may be one hundred or may be several hundred. These names are arranged in different ways, according to the laws of the different states. Under the Massachusetts law, which is considered the best by reformers, the names of candidates for each office are arranged alphabetically on a “blanket” ballot, as it is called from its size, and the elector places a mark opposite the names of such candidates as he may wish to vote for. Other states, New York for example, have the blanket system, but the names of the candidates are arranged in party columns. Still other states allow the grouping on one ballot of all the candidates of a single party, and there would be therefore as many separate ballots in such states as there were parties in the field.

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The qualifications for voting, while varying in the different states in details, are in their main features the same throughout the Union. A residence in the state is required of from three months to two years. Residence is also necessary, but for a shorter period, in the county, city or town, or voting precinct. A few states require the payment of a poll tax. Some require that the voter shall be able to read and understand the Constitution. This latter qualification has been introduced into several of the Southern states, partly at least to disqualify the ignorant coloured voters. In all, or practically all, the states idiots, convicts and the insane are disqualified; in some states paupers; in some of the Western states the Chinese. In some states women are allowed to vote on certain questions, or for the candidates for certain offices, especially school officials; and in four of the Western states women have the same rights of suffrage as men. The number of those who are qualified to vote, but do not avail themselves of the right, varies greatly in the different states and according to the interest taken in the election. As a general rule, but subject to exceptions, the national elections call out the largest number, the state elections next, and the local elections the smallest number of voters. In an exciting national election between 80 and 90% of the qualified voters actually vote, a proportion considerably greater than in Great Britain or Germany.

The tendency of recent years has been towards a decrease both in the number and in the frequency of elections. A president and vice-president are voted for every fourth year, in the years divisible by four, on the first Tuesday following the first Monday of November. Members of the national House of Representatives are chosen for two years on the even-numbered years. State and local elections take place in accordance with state laws, and may or may not be on the same day as the national elections. Originally the rule was for the states to hold annual elections; in fact, so strongly did the feeling prevail of the need in a democratic country for frequent elections, that the maxim “where annual elections end, tyranny begins,” became a political proverb. But opinion gradually changed even in the older or Eastern states, and in 1909 Massachusetts and Rhode Island were the only states in the Union holding annual elections for governor and both houses of the state legislature. In the Western states especially state officers are chosen for longer terms—in the case of the governor often for four years—and the number of elections has correspondingly decreased. Another cause of the decrease in the number of elections is the growing practice of holding all the elections of any year on one and the same day. Before the Civil War Pennsylvania held its state elections several months before the national elections. Ohio and Indiana, until 1885 and 1881 respectively, held their state elections early in October. Maine, Vermont and Arkansas keep to September. The selection of one day in the year for all elections held in that year has resulted in a considerable decrease in the total number.

Another tendency of recent years, but not so pronounced, is to hold local elections in what is known as the “off” year; that is, on the odd-numbered year, when no national election is held. The object of this reform is to encourage independent voting. The average American citizen is only too prone to carry his national political predilections into local elections, and to vote for the local nominees of his party, without regard to the question of fitness of candidates and the fundamental difference of issues involved. This tendency to vote the entire party ticket is the more pronounced because under the system of voting in use in many of the states all the candidates of the party are arranged on one ticket, and it is much easier to vote a straight or unaltered ticket than to change or “scratch” it. Again, the voter, especially the ignorant one, refrains from scratching his ticket, lest in some way he should fail to comply with the technicalities of the law and his vote be lost. On the other hand, if local elections are held on the “off” or odd year, and there be no national or state candidates, the voter feels much more free to select only those candidates whom he considers best qualified for the various offices.

On the important question of the purity of elections it is difficult to speak with precision. In many of the states, especially those with an enlightened public spirit, such as most of the New England states and many of the North-Western, the elections are fairly conducted, there being no intimidation at all, little or no bribery, and an honest count. It can safely be said that through the Union as a whole the tendency of recent years has been decidedly towards greater honesty of elections. This is owing to a number of causes: (1) The selection of a single day for all elections, and the consequent immense number voting on that day. Some years ago, when for instance the Ohio and Indiana elections were held a few weeks before the general election, each party strained every nerve to carry them, for the sake of prestige and the influence on other states. In fact, presidential elections were often felt to turn on the result in these early voting states, and the party managers were none too scrupulous in the means employed to carry them. Bribery has decreased in such states since the change of election day to that of the rest of the country. (2) The enactment in most of the states of the Australian or secret ballot (*q.v.*) laws. These have led to the secrecy of the ballot, and hence to a greater or less extent have prevented intimidation and bribery. (3) Educational or other such test, more particularly in the Southern states, the object of which is to exclude the coloured, and especially the ignorant coloured, voters from the polls. In those southern states in which the coloured vote was large, and still more in those in which it was the majority, it was felt among the whites that intimidation or ballot-box stuffing was justified by the necessity of white supremacy. With the elimination of the coloured vote by educational or other tests the honesty of elections has increased. (4) The enactment of new and more stringent registration laws. Under these laws only those persons are allowed to vote whose names have been placed on the rolls a certain number of days or months before election. These rolls are open to public inspection, and the names may be challenged at the polls, and “colonization” or repeating is therefore almost impossible. (5) The reform of the civil service and the gradual elimination of the vicious principle of "to the victors belong the spoils." With the reform of the civil service elections become less a scramble for office and more a contest of political or economic principle. They bring into the field, therefore, a better class of candidates. (6) The enactment in a number of states of various other laws for the prevention of corrupt practices, for the publication of campaign expenses, and for the prohibition of party workers from coming within a certain specified distance of the polls. In the state of Massachusetts, for instance, an act passed in 1892, and subsequently amended, provides that political committees shall file a full statement, duly sworn to, of all campaign expenditures made by them. The act applies to all public elections except that of town officers, and also covers nominations by caucuses and conventions as well. Apart from his personal expenses such as postage, travelling expenses, &c., a candidate is prohibited from spending anything himself to promote either his nomination or his election, but he is allowed to contribute to the treasury of the political committee. The law places no limit on the amount that these committees may spend. The reform sought by the law is thorough publicity, and not only are details of receipts and expenditures to be published, but the names of contributors and the amount of their contributions. In the state of New York the act which seeks to prevent corrupt practices relies in like manner on the efficacy of publicity, but it is less effective than the Massachusetts law in that it provides simply for the filing by the candidates themselves of sworn statements of their own expenses. There is nothing to prevent their contributing to political committees, and the financial methods and the amounts expended by such committees are not made public. But behind all these causes that have led to more honest elections lies the still greater one of a healthier public spirit. In the reaction following the Civil War all reforms halted. In recent years, however, a new and healthier interest has sprung up in things political; and one result of this improved civic spirit is seen in the various laws for purification of elections. It may now be safely affirmed that in the majority of states the elections are honestly conducted; that intimidation, bribery, stuffing of the ballot boxes or other forms of corruption, when they exist, are owing in large measure to temporary or local causes; and that the tendency of recent years has been towards a decrease in all forms of corruption. 172

The expenses connected with elections, such as the renting and preparing of the polling-places, the payment of the clerks and other officers who conduct the elections and count the vote, are borne by the community. A candidate therefore is not, as far as the law is concerned, liable to any expense whatever. As a matter of fact he does commonly contribute to the party treasury, though in the case of certain candidates, particularly those for the presidency and for judicial offices, financial contributions are not general. The amount of a candidate's contribution varies greatly, according to the office sought, the state in which he lives, and his private wealth. On one occasion, in a district in New York, a candidate for Congress is credibly believed to have spent at one election $50,000. On the other hand, in a Congressional election in a certain district in Massachusetts, the only expenditure of one of the candidates was for the two-cent stamp placed on his letter of acceptance. No estimate of the average amount expended can be made. It is, however, the conclusion of Mr Bryce, in his *American Commonwealth*, that as a rule a seat in Congress costs the candidate less than a seat for a county division in the House of Commons. (See also Ballot.)

ELECTION, in English law, the obligation imposed upon a party by courts of equity to choose between two inconsistent or alternative rights or claims in cases where there is a clear intention of the person from whom he derives one that he should not enjoy both. Thus a testator died seized of property in fee simple and in fee tail—he had two daughters, and devised the fee simple property to one and the entailed property to the other; the first one claimed to have her share of the entailed property as coparcener and also to retain the benefit she took under the will. It was held that she was put to her election whether she would take under the will and renounce her claim to the entailed property or take against the will, in which case she must renounce the benefits she took under the will in so far as was necessary to compensate her sister. As the essence of the doctrine is compensation, a person electing against a document does not lose all his rights under it, but the court will sequester so much only of the benefit intended for him as will compensate the persons disappointed by his election. For the same reason it is necessary that there should be a free and disposable fund passing by the instrument from which compensation can be made in the event of election against the will. If, therefore, a man having a special power of appointment appoint the fund equally between two persons, one being an object of the power and the other not an object, no question of election arises, but the appointment to the person not an object is bad.

Election, though generally arising in cases of wills, may also arise in the case of a deed. There is, however, a distinction to be observed. In the case of a will a clear intention on the part of the testator that he meant to dispose of property not his own must be shown, and parol evidence is not admissible as to this. In the case of a deed, however, no such intention need be shown, for if a deed confers a benefit and imposes a liability on the same person he cannot be allowed to accept the one and reject the other, but this must be distinguished from cases where two separate gifts are given to a person, one beneficial and the other onerous. In such a case no question of election arises and he may take the one and reject the other, unless, indeed, there are words used which make the one conditional on the acceptance of the other.

Election is either express, *e.g.* by deed, or implied; in the latter case it is often a question of considerable difficulty whether there has in fact been an election or not; each case must depend upon the particular circumstances, but quite generally it may be said that the person who has elected must have been capable of electing, aware of the existence of the doctrine of election, and have had the opportunity of satisfying himself of the relative value of the properties between which he has elected. In the case of infants the court will sometimes elect after an inquiry as to which course is the most advantageous, or if there is no immediate urgency, will allow the matter to stand over till the infant attains his majority. In the cases of married women and lunatics the courts will exercise the right for them. It sometimes happens that the parties have so dealt with the property that it would be inequitable to disturb it; in such cases the court will not interfere in order to allow of election.

ELECTORAL COMMISSION, in United States history, a commission created to settle the disputed presidential election of 1876. In this election Samuel J. Tilden, the Democratic candidate, received 184 uncontested electoral votes, and Rutherford B. Hayes, the Republican candidate, 163.[1] The states of Florida, Louisiana, Oregon and South Carolina, with a total of 22 votes, each sent in two sets of electoral ballots,[2] and from each of these states except Oregon one set gave the whole vote to Tilden and the other gave the whole vote to Hayes. From Oregon one set of ballots gave the three electoral votes of the state to Hayes; the other gave two votes to Hayes and one to Tilden.

The election of a president is a complex proceeding, the method being indicated partly in the Constitution, and being partly left to Congress and partly to the states. The manner of selecting the electors is left to state law; the electoral ballots are sent to the president of the Senate, who "shall, in the presence of the Senate and House of Representatives, open all certificates, and the votes shall then be counted." Concerning this provision many questions of vital importance arose in 1876: Did the president of the Senate count the votes, the houses being mere witnesses; or did the houses count them, the president's duties being merely ministerial? Did counting imply the determination of what should be counted, or was it a mere arithmetical process; that is, did the Constitution itself afford a method of settling disputed returns, or was this left to legislation by Congress? Might Congress or an officer of the Senate go behind a state's certificate and review the acts of its certifying officials? Might it go further and examine into the choice of electors? And if it had such powers, might it delegate them to a commission? As regards the procedure of Congress, it seems that, although in early years the president of the Senate not only performed or overlooked the electoral count but also exercised discretion in some matters very important in 1876, Congress early began to assert power, and, at least from 1821 onward, controlled the count, claiming complete power. The fact, however, that the Senate in 1876 was controlled by the Republicans and the House by the Democrats, lessened the chances of any harmonious settlement of these questions by Congress. The country seemed on the verge of civil war. Hence it was that by an act of the 29th of January 1877, Congress created the Electoral Commission to pass upon the contested returns, giving it "the same powers, if any" possessed by itself in the premises, the decisions to stand unless rejected by the two houses separately. The commission was composed of five Democratic and five Republican Congressmen, two justices of the Supreme Court of either party, and a fifth justice chosen by these four. As its members of the commission the Senate chose G.F. Edmunds of Vermont, O.P. Morton of Indiana, and F.T. Frelinghuysen of New Jersey (Republicans); and A.G. Thurman of Ohio and T.F. Bayard of Delaware (Democrats). The House chose Henry B. Payne of Ohio, Eppa Hunton of Virginia, and Josiah G. Abbott of Massachusetts (Democrats); and George F. Hoar of Massachusetts and James A. Garfield of Ohio (Republicans). The Republican judges were William Strong and Samuel F. Miller; the Democratic, Nathan Clifford and Stephen J. Field. These four chose as the fifteenth member Justice Joseph P. Bradley, a Republican but the only member not selected avowedly as a partisan. As counsel for the 173
Democratic candidate there appeared before the commission at different times Charles O'Conor of New York, Jeremiah S. Black of Pennsylvania, Lyman Trumbull of Illinois, R.T. Merrick of the District of Columbia, Ashbel Green of New Jersey, Matthew H. Carpenter of Wisconsin, George Hoadley of Ohio, and W.C. Whitney of New York. W.M. Evarts and E.W. Stoughton of New York and Samuel Shellabarger and Stanley Matthews of Ohio appeared regularly in behalf of Mr Hayes.

The popular vote seemed to indicate that Hayes had carried South Carolina and Oregon, and Tilden Florida and Louisiana. It was evident, however, that Hayes could secure the 185 votes necessary to elect only by gaining every disputed ballot. As the choice of Republican electors in Louisiana had been accomplished by the rejection of several thousand Democratic votes by a Republican returning board, the Democrats insisted that the commission should go behind the returns and correct injustice; the Republicans declared that the state's action was final, and that to go behind the returns would be invading its sovereignty. When this matter came before the commission it virtually accepted the Republican contention, ruling that it could not go behind the returns except on the superficial issues of manifest fraud therein or the eligibility of electors to their office under the Constitution; that is, it could not investigate antecedents of fraud or misconduct of state officials in the results certified. All vital questions were settled by the votes of eight Republicans and seven Democrats; and as the Republican Senate would never concur with the Democratic House in overriding the decisions, all the disputed votes were awarded to Mr Hayes, who therefore was declared elected.

The strictly partisan votes of the commission and the adoption by prominent Democrats and Republicans, both within and without the commission, of an attitude toward states-rights principles quite inconsistent with party tenets and tendencies, have given rise to much severe criticism. The Democrats and the country, however, quietly accepted the decision. The judgments underlying it were two: (1) That Congress rightly claimed the power to settle such contests within the limits set; (2) that, as Justice Miller said regarding these limits, the people had never at any time intended to give to Congress the power, by naming the electors, to "decide who are to be the president and vice-president of the United States."

There is no doubt that Mr Tilden was morally entitled to the presidency, and the correction of the Louisiana frauds would certainly have given satisfaction then and increasing satisfaction later, in the retrospect, to the country. The commission might probably have corrected the frauds without exceeding its Congressional precedents. Nevertheless, the principles of its decisions must be recognized by all save ultra-nationalists as truer to the spirit of the Constitution and promising more for the good of the country than would have been the principles necessary to a contrary decision.

By an act of the 3rd of February 1887 the electoral procedure is regulated in great detail. Under this act determination by a state of electoral disputes is conclusive, subject to certain formalities that guarantee definite action and accurate certification. These formalities constitute “regularity,” and are in all cases judgable by Congress. When Congress is forced by the lack or evident inconclusiveness of state action, or by conflicting state action, to decide disputes, votes are lost unless both houses concur.

Authorities.—J.F. Rhodes, *History of the United States*, vol. 7, covering 1872-1877 (New York, 1906); P.L. Haworth, *The Hayes-Tilden disputed Presidential Election of 1876* (Cleveland, 1906); J.W. Burgess, *Political Science Quarterly*, vol. 3 (1888), pp. 633-653, “The Law of the Electoral Count”; and for the sources. Senate Miscellaneous Document No. 5 (vol. 1), and House Miscel. Doc. No. 13 (vol. 2), 44 Congress, 2 Session, —*Count of the Electoral Vote. Proceedings of Congress and Electoral Commission*,—the latter identical with *Congressional Record*, vol. 5, pt. 4, 44 Cong., 2 Session; also about twenty volumes of evidence on the state elections involved. The volume called *The Presidential Counts* (New York, 1877) was compiled by Mr. Tilden and his secretary.

1 The election of a vice-president was, of course, involved also. William A. Wheeler was the Republican candidate, and Thomas A. Hendricks the Democratic.

2 A second set of electoral ballots had also been sent in from Vermont, where Hayes had received a popular majority vote of 24,000. As these ballots had been transmitted in an irregular manner, the president of the Senate refused to receive them, and was sustained in this action by the upper House.

ELECTORS (Ger. *Kurfürsten*, from *Küren*, O.H.G. *kiosan*, choose, elect, and *Fürst*, prince), a body of German princes, originally seven in number, with whom rested the election of the German king, from the 13th until the beginning of the 19th century. The German kings, from the time of Henry the Fowler (919-936) till the middle of the 13th century, succeeded to their position partly by heredity, and partly by election. Primitive Germanic practice had emphasized the element of heredity. *Reges ex nobilitate sumunt*: the man whom a German tribe recognized as its king must be in the line of hereditary descent from Woden; and therefore the genealogical trees of early Teutonic kings (as, for instance, in England those of the Kentish and West Saxon sovereigns) are carefully constructed to prove that descent from the god which alone will constitute a proper title for his descendants. Even from the first, however, there had been some opening for election; for the principle of primogeniture was not observed, and there might be several competing candidates, all of the true Woden stock. One of these competing candidates would have to be recognized (as the Anglo-Saxons said, *geceosan*); and to this limited extent Teutonic kings may be termed elective from the very first. In the other nations of western Europe this element of election dwindled, and the principle of heredity alone received legal recognition; in medieval Germany, on the contrary, the principle of heredity, while still exercising an inevitable natural force, sank formally into the background, and legal recognition was finally given to the elective principle. *De facto*, therefore, the principle of heredity exercises in Germany a great influence, an influence never more striking than in the period which follows on the formal recognition of the elective principle, when the Habsburgs (like the Metelli at Rome) *fato imperatores fiunt: de jure*, each monarch owes his accession simply and solely to the vote of an electoral college.

This difference between the German monarchy and the other monarchies of western Europe may be explained by various considerations. Not the least important of these is what seems a pure accident. Whereas the Capetian monarchs, during the three hundred years that followed on the election of Hugh Capet in 987, always left an heir male, and an heir male of full age, the German kings again and again, during the same period, either left a minor to succeed to their throne, or left no issue at all. The principle of heredity began to fail because there were no heirs. Again the strength of tribal feeling in Germany made the monarchy into a prize, which must not be the apanage of any single tribe, but must circulate, as it were, from Franconian to Saxon, from Saxon to Bavarian, from Bavarian to Franconian, from Franconian to Swabian; while the growing power of the baronage, and its habit of erecting anti-kings to emphasize its opposition to the crown (as, for instance, in the reign of Henry IV.), coalesced with and gave new force to the action of tribal feeling. Lastly, the fact that the German kings were also Roman emperors finally and irretrievably consolidated the growing tendency towards the elective principle. The principle of heredity had never held any great sway under the ancient Roman Empire (see under Emperor); and the medieval Empire, instituted as it was by the papacy, came definitely under the influence of ecclesiastical prepossessions in favour of election. The church had substituted for that descent from Woden, which had elevated the old pagan kings to their thrones, the conception that the monarch derived his crown from the choice of God, after the manner of Saul; and the theoretical choice of God was readily turned into the actual choice of the church, or, at any rate, of the general body of churchmen. If an ordinary king is thus regarded by the church as essentially elected, much more will the emperor, connected as he is with the church as one of its officers, be held to be also elected; and as a bishop is chosen by the chapter of his diocese, so, it will be thought, must the emperor be chosen by some corresponding body in his empire. Heredity might be tolerated in a mere matter of kingship: the precious trust of imperial power could not be allowed to descend according to the accidents of family succession. To Otto of Freising (*Gesta Frid.* ii. 1) it is already a point of right vindicated for itself by the 174
excellency of the Roman Empire, as a matter of singular prerogative, that it should not descend *per sanguinis propaginem, sed per principum electionem*.

The accessions of Conrad II. (see Wipo, *Vita Cuonradi*, c. 1-2), of Lothair II. (see *Narratio de electione Lotharii*, M.G.H. *Scriptt.* xii. p. 510), of Conrad III. (see Otto of Freising, *Chronicon*, vii. 22) and of Frederick I. (see Otto of Freising, *Gesta Frid.* ii. 1) had all been marked by an element, more or less pronounced, of election. That element is perhaps most considerable in the case of Lothair, who had no rights of heredity to urge. Here we read of ten princes being selected from the princes of the various duchies, to whose choice the rest promise to assent, and of these ten selecting three candidates, one of whom, Lothair, is finally chosen (apparently by the whole assembly) in a somewhat tumultuary fashion. In this case the electoral assembly would seem to be, in the last resort, the whole diet of all the princes. But a *de facto* pre-eminence in the act of election is already, during the 12th century, enjoyed by the three Rhenish archbishops, probably because of the part they afterwards played at the coronation, and also by the dukes of the great duchies—possibly because of the part they too played, as vested for the time with the great offices of the household, at the coronation feast.[1] Thus at the election of Lothair it is the archbishop of Mainz who conducts the proceedings; and the election is not held to be final until the duke of Bavaria has given his assent. The fact is that, votes being weighed by quality as well as by quantity (see DIET), the votes of the archbishops and dukes, which would first be taken, would of themselves, if unanimous, decide the election. To prevent tumultuary elections, it was well that the election should be left exclusively with these great dignitaries; and this is what, by the middle of the 13th century, had eventually been done.

The chaos of the interregnum from 1198 to 1212 showed the way for the new departure; the chaos of the great interregnum (1250-1273) led to its being finally taken. The decay of the great duchies, and the narrowing of the class of princes into a close corporation, some of whose members were the equals of the old dukes in power, introduced difficulties and doubts into the practice of election which had been used in the 12th century. The contested election of the interregnum of 1198-1212 brought these difficulties and doubts into strong relief. The famous bull of Innocent III. (*Venerabilem*), in which he decided for Otto IV. against Philip of Swabia, on the ground that, though he had fewer votes than Philip, he had a majority of the votes of those *ad quos principaliter spectat electio*, made it almost imperative that there should be some definition of these principal electors. The most famous attempt at such a definition is that of the *Sachsenspiegel*, which was followed, or combated, by many other writers in the first half of the 13th century. Eventually the contested election of 1257 brought light and definition. Here we find seven potentates acting—the same seven whom the Golden Bull recognizes in 1356; and we find these seven described in an official letter to the pope, as *principes vocem in hujusmodi electione habentes, qui sunt septem numero*. The doctrine thus enunciated was at once received. The pope acknowledged it in two bulls (1263); a cardinal, in a commentary on the bull *Venerabilem* of Innocent III., recognized it about the same time; and the erection of statues of the seven electors at Aix-la-Chapelle gave the doctrine a visible and outward expression.

By the date of the election of Rudolph of Habsburg (1273) the seven electors may be regarded as a definite body, with an acknowledged right. But the definition and the acknowledgment were still imperfect. (1) The composition of the electoral body was uncertain in two respects. The duke of Bavaria claimed as his right the electoral vote of the king of Bohemia; and the practice of *partitio* in electoral families tended to raise further difficulties about the exercise of the vote. The Golden Bull of 1356 settled both these questions. Bohemia (of which Charles IV., the author of the Golden Bull, was himself the king) was assigned the electoral vote in preference to Bavaria; and a provision annexing the electoral vote to a definite territory, declaring that territory indivisible, and regulating its descent by the rule of primogeniture instead of partition, swept away the old difficulties which the custom of partition had raised. After 1356 the seven electors are regularly the three Rhenish archbishops, Mainz, Cologne and Trier, and four lay magnates, the palatine of the Rhine, the duke of Saxony, the margrave of Brandenburg, and the king of Bohemia; the three former being vested with the three archchancellorships, and the four latter with the four offices of the royal household (see HOUSEHOLD). (2) The rights of the seven electors, in their collective capacity as an electoral college, were a matter of dispute with the papacy. The result of the election, whether made, as at first, by the princes generally or, as after 1257, by the seven electors exclusively, was in itself simply the creation of a German king—an *electio in regem*. But since 962 the German king was also, after coronation by the pope, Roman emperor. Therefore the election had a double result: the man elected was not only *electus in regem*, but also *promovendus ad imperium*. The difficulty was to define the meaning of the term *promovendus*. Was the king elect *inevitably* to become emperor? or did the *promotio* only follow at the discretion of the pope, if he thought the king elect fit for promotion? and if so, to what extent, and according to what standard, did the pope judge of such fitness? Innocent III. had already claimed, in the bull *Venerabilem*, (1) that the electors derived their power of election, so far as it made an emperor, from the Holy See (which had originally "translated" the Empire from the East to the West), and (2) that the papacy had a *jus et auctoritas examinandi personam electam in regem et promovendam ad imperium*. The latter claim he had based on the fact that he anointed, consecrated and crowned the emperor—in other words, that he gave a spiritual office according to spiritual methods, which entitled him to inquire into the fitness of the recipient of that office, as a bishop inquires into the fitness of a candidate for ordination. Innocent had put forward this claim as a ground for deciding between competing candidates: Boniface VIII. pressed the claim against Albert I. in 1298, even though his election was unanimous; while John XXII. exercised it in its harshest form, when in 1324 he ex-communicated Louis IV. for using the title and exerting the rights even of king without previous papal confirmation. This action ultimately led to a protest from the electors themselves, whose right of election would have become practically meaningless, if such assumptions had been tolerated. A meeting of the electors (*Kurverein*) at Rense in 1338 declared (and the declaration was reaffirmed by a diet at Frankfort in the same year) that *postquam aliquis eligitur in Imperatorem sive Regem ab Electoribus Imperii concorditer, vel majori parte eorundem, statim ex sola electione est Rex verus et Imperator Romanus censendus ... nec Papae sive Sedis Apostolicae ... approbatione ... indiget*. The doctrine thus positively affirmed at Rense is negatively reaffirmed in the Golden Bull, in which a significant silence is maintained in regard to papal rights. But the doctrine was not in practice followed: Sigismund himself did not venture to dispense with papal approbation.

By the end of the 14th century the position of the electors, both individually and as a corporate body, had become definite and precise. Individually, they were distinguished from all other princes, as we have seen, by the indivisibility of their territories and by the custom of primogeniture which secured that indivisibility; and they were still further distinguished by the fact that their person, like that of the emperor himself, was protected by the law of treason, while their territories were only subject to the jurisdiction of their own courts. They were independent territorial sovereigns; and their position was at once the envy and the ideal of the other princes of Germany. Such had been the policy of Charles IV.; and thus had he, in the Golden Bull, sought to magnify the seven electors, and himself as one of the seven, in his capacity of king of Bohemia, even at the expense of the Empire, and of himself in his capacity of emperor. Powerful as they were, however, in their individual capacity, the electors showed themselves no less powerful as a corporate body. As such a corporate body, they may be considered from three different points of view, and as acting in three different capacities. They are an electoral body, choosing each successive emperor; they are one of the three colleges of the imperial diet (see DIET); and they are also an electoral union (*Kurfürstenverein*), acting as a separate and independent political organ even after the election, and during the reign, of the monarch. It was in this last capacity that they had met at Rense in 1338; and in the same capacity they acted repeatedly during the 15th century. According to the Golden Bull, such meetings were to be annual, and their deliberations were to concern "the safety of the Empire and the world." Annual they never were; but occasionally they became of great importance. In 1424, during the attempt at reform occasioned by the failure of German arms against the Hussites, the *Kurfürstenverein* acted, or at least it claimed to act, as the predominant partner in a duumvirate, in which the unsuccessful Sigismund was relegated to a secondary position. During the long reign of Frederick III.—a reign in which the interests of Austria were cherished, and the welfare of the Empire neglected, by that apathetic yet tenacious emperor—the electors once more attempted, in the year 1453, to erect a new central government in place of the emperor, a government which, if not conducted by themselves directly in their capacity of a *Kurfürstenverein*, should at any rate be under their influence and control. So, they hoped, Germany might be able to make head against that papal aggression, to which Frederick had yielded, and to take a leading part in that crusade against the Turks, which he had neglected. Like the previous attempt at reform during the Hussite wars, the scheme came to nothing; the forces of disunion in Germany were too strong for any central government, whether monarchical and controlled by the emperor, or oligarchical and controlled by the electors. But a final attempt, the most strenuous of all, was made in the reign of Maximilian I., and under the influence of Bertold, elector and archbishop of Mainz. The council of 1500, in which the electors (with the exception of the king of Bohemia) were to have sat, and which would have been under their control, represents the last effective attempt at a real *Reichsregiment*. Inevitably, however, it shipwrecked on the opposition of Maximilian; and though the attempt was again made between 1521 and 1530, the idea of a real central government under the control of the electors perished, and the development of local administration by the circle took its place.

In the course of the 16th century a new right came to be exercised by the electors. As an electoral body (that is to say, in the first of the three capacities distinguished above), they claimed, at the election of Charles V. in 1519 and at subsequent elections, to impose conditions on the elected monarch, and to determine the terms on which he should exercise his office in the course of his reign. This *Wahlcapitulation*, similar to the *Pacta Conventa* which limited the elected kings of Poland, was left by the diet to the discretion of the electors, though after the treaty of Westphalia an attempt was made, with some little success,[2] to turn the capitulation into a matter of legislative enactment by the diet. From this time onwards the only fact of importance in the history of the electors is the change which took place in the composition of their body during the 17th and 18th centuries. From the Golden Bull to the treaty of Westphalia (1356-1648) the composition of the electoral body had remained unchanged. In 1623, however, in the course of the Thirty Years' War, the vote of the count palatine of the Rhine had been transferred to the duke of Bavaria; and at the treaty of Westphalia the vote, with the office of imperial butler which it carried, was left to Bavaria, while an eighth vote, along with the new office of imperial treasurer, was created for the count palatine. In 1708 a ninth vote, along with the office of imperial standard-bearer, was created for Hanover; while finally, in 1778, the vote of Bavaria and the office of imperial butler returned to the counts palatine, as heirs of the duchy, on the extinction of the ducal line, while the new vote created for the Palatinate in 1648, with the office of imperial treasurer, was transferred to Brunswick-Lüneburg (Hanover) in lieu of the one which this house already held. In 1806, on the dissolution of the Holy Roman Empire, the electors ceased to exist.

LITERATURE.—T. Lindner, *Die deutschen Königswahlen und die Entstehung des Kurfürstentums* (1893), and *Der Hergang bei den deutschen Königswahlen* (1899); R. Kirchhöfer, *Zur Entstehung des Kurkollegiums* (1893); W. Maurenbrecher, *Geschichte der deutschen Königswahlen* (1889); and G. Blondel, *Étude sur Frédéric II*, p. 27 sqq. See also J. Bryce, *Holy Roman Empire* (edition of 1904), c. ix.; and R. Schröder, *Lehrbuch der deutschen Rechtsgeschichte*, pp. 471-481 and 819-820.

(E. BR.)

1 This is the view of the *Sachsenspiegel*, and also of Albert of Stade (quoted in Schröder, p. 476, n. 27): "Palatinus eligit, quia dapifer est; dux Saxoniae, quia marescalcus," &c. Schröder points out (p. 479, n. 45) that "participation in the coronation feast is an express recognition of the king"; and those who are to discharge their office in the one must have had a prominent voice in the other.

2 See Schröder's *Lehrbuch der deutschen Rechtsgeschichte*, p. 820.

ELECTRA (Ἠλέκτρα), "the bright one," in Greek mythology. (1) One of the seven Pleiades, daughter of Atlas and Pleïone. She is closely connected with the old constellation worship and the religion of Samothrace, the chief seat of the Cabeiri (*q.v.*), where she was generally supposed to dwell. By Zeus she was the mother of Dardanus, Iasion (or Eëtion), and Harmonia; but in the Italian tradition, which represented Italy as the original home of the Trojans, Dardanus was her son by a king of Italy named Corythus. After her amour with Zeus, Electra fled to the Palladium as a suppliant, but Athena, enraged that it had been touched by one who was no longer a maiden, flung Electra and the image from heaven to earth, where it was found by Ilus, and taken by him to Ilium; according to another tradition, Electra herself took it to Ilium, and gave it to her son Dardanus (Schol. Eurip. *Phoen.* 1136). In her grief at the destruction of the city she plucked out her hair and was changed into a comet; in another version Electra and her six sisters had been placed among the stars as the Pleiades, and the star which she represented lost its brilliancy after the fall of Troy. Electra's connexion with Samothrace (where she was also called Electryone and Strategis) is shown by the localization of the carrying off of her reputed daughter Harmonia by Cadmus, and by the fact that, according to Athenicon (the author of a work on Samothrace quoted by the scholiast on Apollonius Rhodius i. 917), the Cabeiri were Dardanus and Iasion. The gate Electra at Thebes and the fabulous island Electris were said to have been called after her (Apollodorus iii. 10. 12; Servius on *Aen.* iii. 167, vii. 207, x. 272, *Georg.* i. 138).

(2) Daughter of Agamemnon and Clytaemnestra, sister of Orestes and Iphigeneia. She does not appear in Homer, although according to Xanthus (regarded by some as a fictitious personage), to whom Stesichorus was indebted for much in his *Oresteia*, she was identical with the Homeric Laodice, and was called Electra because she remained so long unmarried (Ἀ-λέκτρα). She was said to have played an important part in the poem of Stesichorus, and subsequently became a favourite figure in tragedy. After the murder of her father on his return from Troy by her mother and Aegisthus, she saved the life of her brother Orestes by sending him out of the country to Strophius, king of Phanote in Phocis, who had him brought up with his own son Pylades. Electra, cruelly ill-treated by Clytaemnestra and her paramour, never loses hope that her brother will return to avenge his father. When grown up, Orestes, in response to frequent messages from his sister, secretly repairs with Pylades to Argos, where he pretends to be a messenger from Strophius bringing the news of the death of Orestes. Being admitted to the palace, he slays both Aegisthus and Clytaemnestra. According to another story (Hyginus, *Fab.* 122), Electra, having received a false report that Orestes and Pylades had been sacrificed to Artemis in Tauris, went to consult the oracle at Delphi. In the meantime Aletes, the son of Aegisthus, seized the throne of Mycenae. Her arrival at Delphi coincided with that of Orestes and Iphigeneia. The same messenger, who had already communicated the false report of the death of Orestes, informed her that he had been slain by Iphigeneia. Electra in her rage seized a burning brand from the altar, intending to blind her sister; but at the critical moment Orestes appeared, recognition took place, and the brother and sister returned to Mycenae. Aletes was slain by Orestes, and Electra became the wife of 176
Pylades. The story of Electra is the subject of the *Choëphori* of Aeschylus, the *Electra* of Sophocles and the *Electra* of Euripides. It is in the Sophoclean play that Electra is most prominent.

There are many variations in the treatment of the legend, for which, as also for a discussion of the modern plays on the subject by Voltaire and Alfieri, see Jebb’s Introduction to his edition of the *Electra* of Sophocles.

ELECTRICAL (or Electrostatic) **MACHINE,** a machine operating by manual or other power for transforming mechanical work into electric energy in the form of electrostatic charges of opposite sign delivered to separate conductors. Electrostatic machines are of two kinds: (1) Frictional, and (2) Influence machines.

Frictional Machines.—A primitive form of frictional electrical machine was constructed about 1663 by Otto von Guericke (1602-1686). It consisted of a globe of sulphur fixed on an axis and rotated by a winch, and it was electrically excited by the friction of warm hands held against it. Sir Isaac Newton appears to have been the first to use a glass globe instead of sulphur (*Optics*, 8th Query). F. Hawksbee in 1709 also used a revolving glass globe. A metal chain resting on the globe served to collect the charge. Later G.M. Bose (1710-1761), of Wittenberg, added the prime conductor, an insulated tube or cylinder supported on silk strings, and J.H. Winkler (1703-1770), professor of physics at Leipzig, substituted a leather cushion for the hand. Andreas Gordon (1712-1751) of Erfurt, a Scotch Benedictine monk, first used a glass cylinder in place of a sphere. Jesse Ramsden (1735-1800) in 1768 constructed his well-known form of plate electrical machine (fig. 1). A glass plate fixed to a wooden or metal shaft is rotated by a winch. It passes between two rubbers made of leather, and is partly covered with two silk aprons which extend over quadrants of its surface. Just below the places where the aprons terminate, the glass is embraced by two insulated metal forks having the sharp points projecting towards the glass, but not quite touching it. The glass is excited positively by friction with the rubbers, and the charge is drawn off by the action of the points which, when acted upon inductively, discharge negative electricity against it. The insulated conductor to which the points are connected therefore becomes positively electrified. The cushions must be connected to earth to remove the negative electricity which accumulates on them. It was found that the machine acted better if the rubbers were covered with bisulphide of tin or with F. von Kienmayer’s amalgam, consisting of one part of zinc, one of tin and two of mercury. The cushions were greased and the amalgam in a state of powder spread over them. Edward Nairne’s electrical machine (1787) consisted of a glass cylinder with two insulated conductors, called prime conductors, on glass legs placed near it. One of these carried the leather exacting cushions and the other the collecting metal points, a silk apron extending over the cylinder from the cushion almost to the points. The rubber was smeared with amalgam. The function of the apron is to prevent the escape of electrification from the glass during its passage from the rubber to the collecting points. Nairne’s machine could give either positive or negative electricity, the first named being collected from the prime conductor carrying the collecting points and the second from the prime conductor carrying the cushion.

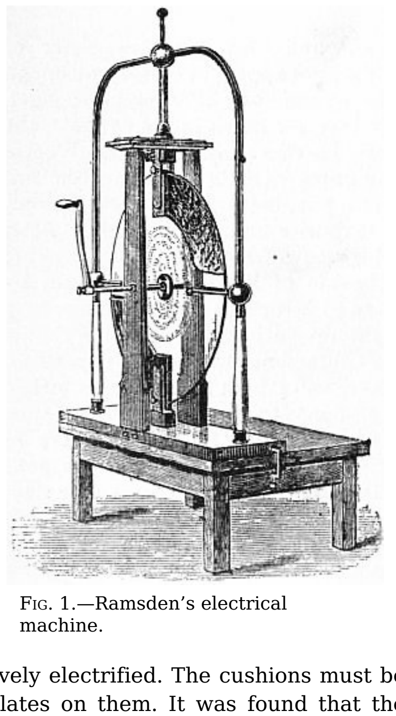

Fig. 1.—Ramsden’s electrical machine.

Influence Machines.—Frictional machines are, however, now

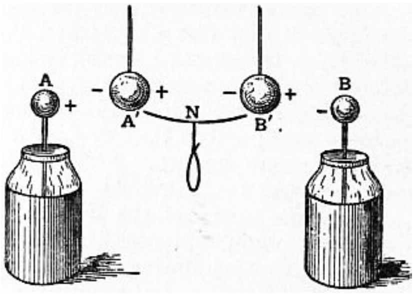

FIG. 2.

quite superseded by the second class of instrument mentioned above, namely, influence machines. These operate by electrostatic induction and convert mechanical work into electrostatic energy by the aid of a small initial charge which is continually being replenished or reinforced. The general principle of all the machines described below will be best understood by considering a simple ideal case. Imagine two Leyden jars with large brass knobs, A and B, to stand on the ground (fig. 2). Let one jar be initially charged with positive electricity on its inner coating and the other with negative, and let both have their outsides connected to earth. Imagine two insulated balls A′ and B′ so held that A′ is near A and B′ is near B. Then the positive charge on A induces two charges on A′, viz.: a negative on the side nearest and a positive on the side most removed. Likewise the negative charge on B induces a positive charge on the side of B′ nearest to it and repels negative electricity to the far side. Next let the balls A′ and B′ be connected together for a moment by a wire N called a neutralizing conductor which is subsequently removed. Then A′ will be left negatively electrified and B′ will be left positively electrified. Suppose that A′ and B′ are then made to change places. To do this we shall have to exert energy to remove A′ against the attraction of A and B′ against the attraction of B. Finally let A′ be brought in contact with B and B′ with A. The ball A′ will give up its charge of negative electricity to the Leyden jar B, and the ball B′ will give up its positive charge to the Leyden jar A. This transfer will take place because the inner coatings of the Leyden jars have greater capacity with respect to the earth than the balls. Hence the charges of the jars will be increased. The balls A′ and B′ are then practically discharged, and the above cycle of operations may be repeated. Hence, however small may be the initial charges of the Leyden jars, by a principle of accumulation resembling that of compound interest, they can be increased as above shown to any degree. If this series of operations be made to depend upon the continuous rotation of a winch or handle, the arrangement constitutes an electrostatic influence machine. The principle therefore somewhat resembles that of the self-exciting dynamo.

Bennet's Doubler.

The first suggestion for a machine of the above kind seems to have grown out of the invention of Volta's electrophorus. Abraham Bennet, the inventor of the gold leaf electroscope, described a doubler or machine for multiplying electric charges (*Phil. Trans.*, 1787).

The principle of this apparatus may be explained thus. Let A and C be two fixed disks, and B a disk which can be brought at will within a very short distance of either A or C. Let us suppose all the plates to be equal, and let the capacities of A and C in presence of B be each equal to p, and the coefficient of induction between A and B, or C and B, be q. Let us also suppose that the plates A and C are so distant from each other that there is no mutual influence, and that p' is the capacity of one of the disks when it stands alone. A small charge Q is communicated to A, and A is insulated, and B, uninsulated, is brought up to it; the charge on B will be—(q/p)Q. B is now uninsulated and brought to face C, which is uninsulated; the charge on C will be $(q/p)^2Q$. C is now insulated and connected with A, which is always insulated. B is then brought to face A and uninsulated, so that the charge on A becomes rQ, where

$$r = \frac{p}{p + p'} \left(1 + \frac{q^2}{p^2} \right).$$

A is now disconnected from C, and here the first operation ends. It is obvious that at the end of n such operations the charge on A will be r^nQ, so that the charge goes on increasing in geometrical progression. If the distance between the disks could be made infinitely small each time, then the multiplier r would be 2, and 177
the charge would be doubled each time. Hence the name of the apparatus.

Erasmus Darwin, B. Wilson, G.C. Bohnenberger and J.C.E. Peclet devised various modifications of Bennet's instrument (see S.P. Thompson, "The Influence Machine from 1788 to 1888," *Journ. Soc. Tel. Eng.*, 1888, 17, p. 569). Bennet's doubler appears to have given a suggestion to William Nicholson (*Phil. Trans.*, 1788, p. 403) of "an instrument which by turning a winch produced the two states of electricity without friction or communication with the earth." This "revolving doubler," according to the description of Professor S.P. Thompson (*loc. cit.*), consists of two fixed plates of brass A and C (fig. 3), each two inches in diameter and separately supported on insulating arms in the same plane, so that a third revolving plate B may pass very near them without touching. A brass ball D two inches in diameter is fixed on the end of the axis that carries the plate B, and is loaded within at one side, so as to act as a counterpoise to the revolving plate B. The axis P N is made of varnished glass, and so are the axes that join the three plates with the brass axis N O. The axis N O passes through the brass piece M, which stands on an insulating pillar of glass, and supports the plates A and C. At one extremity of this axis is the ball D, and the other is connected with a rod of glass, N P, upon which is fixed the handle L, and also the piece G H, which is separately insulated. The pins E, F rise out of the back of the fixed plates A and C, at unequal distances from the axis. The piece K is parallel to G H, and both of them are furnished at their ends with small pieces of flexible wire that they may touch the pins E, F in certain points of their revolution. From the brass piece M there stands out a pin I, to touch against a small flexible wire or spring which projects sideways from the rotating plate B when it comes opposite A. The wires are so adjusted by bending that B, at the moment when it is opposite A, communicates with the ball D, and A communicates with C through GH; and half a revolution later C, when B comes opposite to it, communicates with the ball D through the contact of K with F. In all other positions A, B, C and D are completely disconnected from each other. Nicholson thus described the operation of his machine:—

Nicholson's doubler.

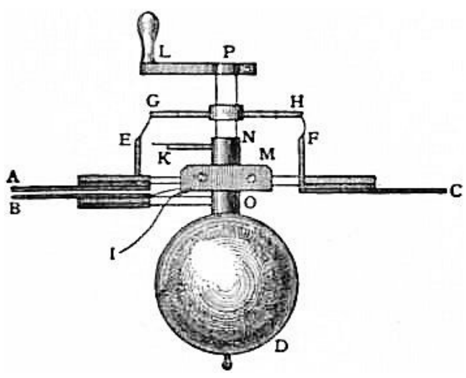

FIG. 3.—Nicholson's Revolving Doubler.

"When the plates A and B are opposite each other, the two fixed plates A and C may be considered as one mass, and the revolving plate B, together with the ball D, will constitute another mass. All the experiments yet made concur to prove that these two masses will not possess the same electric state.... The redundant electricities in the masses under consideration will be unequally distributed; the plate A will have about ninety-nine parts, and the plate C one; and, for the same reason, the revolving plate B will have ninety-nine parts of the opposite electricity, and the ball D one. The rotation, by destroying the contacts, preserves this unequal distribution, and carries B from A to C at the same time that the tail K connects the ball with the plate C. In this situation, the electricity in B acts upon that in C, and produces the contrary state, by virtue of the communication between C and the ball; which last must therefore acquire an electricity of the same kind with that of the revolving plate. But the rotation again destroys the contact and restores B to its first situation opposite A. Here, if we attend to the effect of the whole revolution, we shall find that the electric states of the respective masses have been greatly increased; for the ninety-nine parts in A and B remain, and the one part of electricity in C has been increased so as nearly to compensate ninety-nine parts of the opposite electricity in the revolving plate B, while the communication produced an opposite mutation in the electricity of the ball. A second rotation will, of course, produce a proportional augmentation of these increased quantities; and a continuance of turning will soon bring the intensities to their maximum, which is limited by an explosion between the plates" (*Phil. Trans.*, 1788, p. 405).

Nicholson described also another apparatus, the "spinning condenser," which worked on the same principle. Bennet and Nicholson were followed by T. Cavallo, John Read, Bohnenberger, C.B. Désormes and J.N.P. Hachette and others in the invention of various forms of rotating doubler. A simple and typical form of doubler, devised in 1831 by G. Belli (fig. 4), consisted of two curved metal plates between which revolved a pair of balls carried on an insulating stem. Following the nomenclature usual in connexion with dynamos we may speak of the conductors which carry the initial charges as the field plates, and of the moving conductors on which are induced the charges which are subsequently added to those on the field plates, as the carriers. The wire which connects two armature plates for a moment is the neutralizing conductor. The two curved metal plates constitute the field plates and must have original charges imparted to them of opposite sign. The rotating balls are the carriers, and are connected together for a moment by a wire when in a position to be acted upon inductively by the field plates, thus acquiring charges of opposite sign. The moment after they are separated again. The rotation continuing the ball thus negatively charged is made to give up this charge to that negatively electrified field plate, and the ball positively charged its charge to the positively electrified field plate, by touching little contact springs. In this manner the field plates accumulate charges of opposite sign.

Belli's doubler.

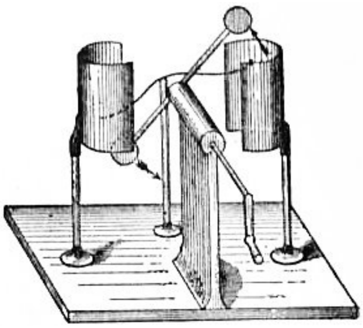

Fig. 4.—Belli's Doubler.

Modern types of influence machine may be said to date from 1860 when C.F. Varley patented a type of influence machine which has been the parent of numerous subsequent forms (*Brit. Pat. Spec.* No. 206 of 1860). In it the field plates were sheets of tin-foil attached to a glass plate (fig. 5). In front of them a disk of ebonite or glass, having carriers of metal fixed to its edge, was rotated by a winch. In the course of their rotation two diametrically opposite carriers touched against the ends of a neutralizing conductor so as to form for a moment one conductor, and the moment afterwards these two carriers were insulated, one carrying away a positive charge and the other a negative. Continuing their rotation, the positively charged carrier gave up its positive charge by touching a little knob attached to the positive field plate, and similarly for the negative charge carrier. In this way the charges on the field plates were continually replenished and reinforced. Varley also constructed a multiple form of influence machine having six rotating disks, each having a number of carriers and rotating between field plates. With this apparatus he obtained sparks 6 in. long, the initial source of electrification being a single Daniell cell.

Varley's machine.

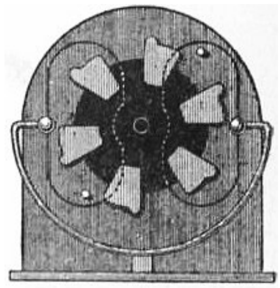

Fig. 5.—Varley's Machine.

Varley was followed by A.J.I. Toepler, who in 1865 constructed an influence machine consisting of two disks fixed on the same shaft and rotating in the same direction. Each disk carried two strips of tin-foil extending nearly over a semi-circle, and there were two field plates, one behind each disk; one of the plates was positively and the other negatively electrified. The carriers which were touched under the influence of the positive field plate passed on and gave up a portion of their negative charge to increase that of the negative field plate; in the same way the carriers which were touched under the influence of the negative field plate sent a part of their charge to augment that of the positive field plate. In this apparatus one of the charging rods communicated with one of the field plates, but the other with the neutralizing brush opposite to the other field plate. Hence one of the field plates would always remain charged when a spark was taken at the transmitting terminals. 178

Toepler machine.

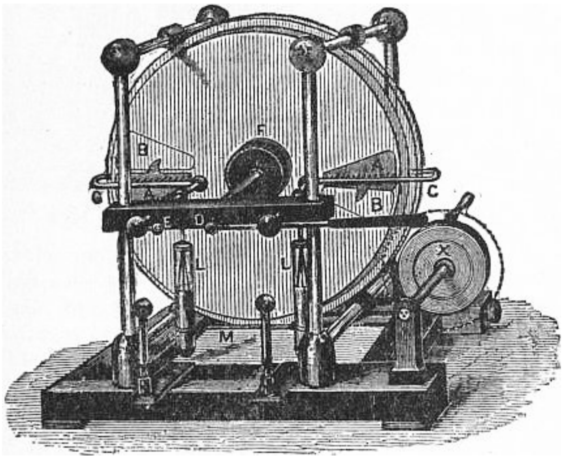

Fig. 6.—Holtz's Machine.

Between 1864 and 1880, W.T.B. Holtz constructed and described a large number of influence machines which were for a long time considered the most advanced development of this type of electrostatic machine.

Holtz machine.

In one form the Holtz machine consisted of a glass disk mounted on a horizontal axis F (fig. 6) which could be made to rotate at a considerable speed by a multiplying gear, part of which is seen at X. Close behind this disk was fixed another vertical disk of glass in which were cut two windows B, B. On the side of the fixed disk next the rotating disk were pasted two sectors of paper A, A, with short blunt points attached to them which projected out into the windows on the side away from the rotating disk. On the other side of the rotating disk were placed two metal combs C, C, which consisted of sharp points set in metal rods and were each connected to one of a pair of discharge balls E, D, the distance between which could be varied. To start the machine the balls were brought in contact, one of the paper armatures electrified, say, with positive electricity, and the disk set in motion. Thereupon very shortly a hissing sound was heard and the machine became harder to turn as if the disk were moving through a resisting medium. After that the discharge balls might be separated a little and a continuous series of sparks or brush discharges would take place between them. If two Leyden jars L, L were hung upon the conductors which supported the combs, with their outer coatings put in connexion with one another by M, a series of strong spark discharges passed between the discharge balls. The action of the machine is as follows: Suppose one paper armature to be charged positively, it acts by induction on the right hand comb, causing negative electricity to issue from the comb points upon the glass revolving disk; at the same time the positive electricity passes through the closed discharge circuit to the left comb and issues from its teeth upon the part of the glass disk at the opposite end of the diameter. This positive electricity electrifies the left paper armature by induction, positive electricity issuing from the blunt point upon the side farthest from the rotating disk. The charges thus deposited on the glass disk are carried round so that the upper half is electrified negatively on both sides and the lower half positively on both sides, the sign of the electrification being reversed as the disk passes between the combs and the armature by discharges issuing from them respectively. If it were not for leakage in various ways, the electrification would go on everywhere increasing, but in practice a stationary state is soon attained. Holtz's machine is very uncertain in its action in a moist climate, and has generally to be enclosed in a chamber in which the air is kept artificially dry.

Voss's machine.

Robert Voss, a Berlin instrument maker, in 1880 devised a form of machine in which he claimed that the principles of Toepler and Holtz were combined. On a rotating glass or ebonite disk were placed carriers of tin-foil or metal buttons against which neutralizing brushes touched. This armature plate revolved in front of a field plate carrying two pieces of tin-foil backed up by larger pieces of varnished paper. The studs on the armature plate were charged inductively by being connected for a moment by a neutralizing wire as they passed in front of the field plates, and then gave up their charges partly to renew the field charges and partly to collecting combs connected to discharge balls. In general design and construction, the manner of moving the rotating plate and in the use of the two Leyden jars in connexion with the discharge balls, Voss borrowed his ideas from Holtz.

Wimshurst machine.

All the above described machines, however, have been thrown into the shade by the invention of a greatly improved type of influence machine first constructed by James Wimshurst about 1878. Two glass disks are mounted on two shafts in such a manner that, by means of two belts and pulleys worked from a winch shaft, the disks can be rotated rapidly in opposite directions close to each other (fig. 7). These glass disks carry on them a certain number (not less than 16 or 20) tin-foil carriers which may or may not have brass buttons upon them. The glass plates are well varnished, and the carriers are placed on the outer sides of the two glass plates. As therefore the disks revolve, these carriers travel in opposite directions, coming at intervals in opposition to each other. Each upright bearing carrying the shafts of the revolving disks also carries a neutralizing conductor or wire ending in a little brush of gilt thread. The neutralizing conductors for each disk are placed at right angles to each other. In addition there are collecting combs which occupy an intermediate position and have sharp points projecting inwards, and coming near to but not touching the carriers. These combs on opposite sides are connected respectively to the inner coatings of two Leyden jars whose outer coatings are in connexion with one another.

The operation of the machine is as follows: Let us suppose that one of the studs on the back plate is positively electrified and one at the opposite end of a diameter is negatively electrified, and that at that moment two corresponding studs on the front plate passing opposite to these back studs are momentarily connected together by the neutralizing wire belonging to the front plate.

The positive stud on the back plate will act inductively on the front stud and charge it negatively, and similarly for the other stud, and as the rotation continues these charged studs will pass round and give up most of their charge through the combs to the Leyden jars. The moment, however, a pair of studs on the front plate are charged, they act as field plates to studs on the back plate which are passing at the moment, provided these last are connected by the back neutralizing wire. After a few revolutions of the disks half the studs on the front plate at any moment are charged negatively and half positively and the same on the back plate, the neutralizing wires forming the boundary between the positively and negatively charged studs. The diagram in fig. 8, taken by permission from S.P. Thompson’s paper (*loc. cit.*), represents a view of the distribution of these charges on the front and back plates respectively. It will be seen that each stud is in turn both a field plate and a carrier having a charge induced on it, and then passing on in turn induces further charges on other studs. Wimshurst constructed numerous very powerful machines of this type, some of them with multiple plates, which operate in almost any climate, and rarely fail to charge themselves and deliver a torrent of sparks between the discharge balls whenever the winch is turned. He also devised an alternating current electrical machine in which the discharge balls were alternately positive and negative. Large Wimshurst multiple plate influence machines are often used instead of induction coils for exciting Röntgen ray tubes in medical work. They give very steady illumination on fluorescent screens.

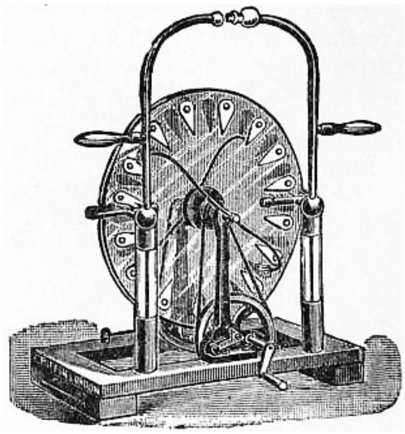

Fig. 7.—Wimshurst’s Machine.

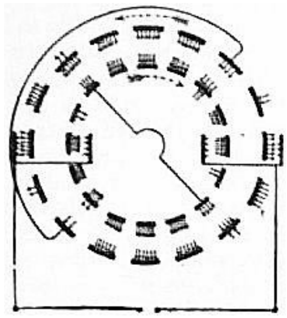

Fig. 8.—Action of the Wimshurst Machine.

In 1900 it was found by F. Tudsbury that if an influence machine is enclosed in a metallic chamber containing compressed air, or better, carbon dioxide, the insulating properties of compressed gases enable a greatly improved effect to be obtained owing to the diminution of the leakage across the plates and from the supports. Hence sparks can be obtained of more than double the length at ordinary atmospheric pressure. In one case a machine with plates 8 in. in diameter which could give sparks 2.5 in. at ordinary pressure gave sparks of 5, 7, and 8 in. as the pressure was raised to 15, 30 and 45 ℔ above the normal atmosphere.

The action of Lord Kelvin’s replenisher (fig. 9) used by him in connexion with his electrometers for maintaining their charge, closely resembles that of Belli’s doubler and will be understood from fig. 9. Lord Kelvin also devised an influence machine, commonly called a “mouse mill,” for electrifying the ink in connexion with his siphon recorder. It was an electrostatic and electromagnetic machine combined, driven by an electric current and producing in turn electrostatic charges of electricity. In connexion with this subject mention must also be made of the water dropping influence machine of the same inventor.[1]

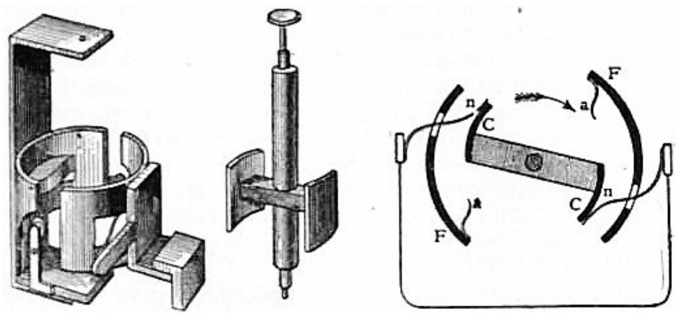

Fig. 9.—Lord Kelvin’s Replenisher.

C, C, Metal carriers fixed to ebonite cross-arm.

F, F, Brass field-plates or conductors.

a, *a*, Receiving springs.

n, *n*, Connecting springs or neutralizing brushes.

The action and efficiency of influence machines have been investigated by F. Rossetti, A. Righi and F.W.G. Kohlrausch. The electromotive force is practically constant no matter what the velocity of the disks, but according to some observers the internal resistance decreases as the velocity increases. Kohlrausch, using a Holtz machine with a plate 16 in. in diameter, found that the current given by it could only electrolyse acidulated water in 40 hours sufficient to liberate one cubic centimetre of mixed gases. E.E.N. Mascart, A. Roiti, and E. Bouchotte have also examined the efficiency and current producing power of influence machines.

Bibliography.—In addition to S.P. Thompson’s valuable paper on influence machines (to which this article is much indebted) and other references given, see J. Clerk Maxwell, *Treatise on Electricity and Magnetism* (2nd ed., Oxford, 1881), vol. i. p. 294; J.D. Everett, *Electricity* (expansion of part iii. of Deschanel’s *Natural Philosophy*) (London, 1901), ch. iv. p. 20; A. Winkelmann, *Handbuch der Physik* (Breslau, 1905), vol. iv. pp. 50-58 (contains a large number of references to original papers); J. Gray, *Electrical Influence Machines, their Development and Modern Forms* (London, 1903).

(J. A. F.)

1 See Lord Kelvin, *Reprint of Papers on Electrostatics and Magnetism* (1872); "Electrophoric Apparatus and Illustrations of Voltaic Theory," p. 319; "On Electric Machines Founded on Induction and Convection," p. 330; "The Reciprocal Electrophorus," p. 337.

ELECTRIC EEL (*Gymnotus electricus*), a member of the family of fishes known as *Gymnotidae*. In spite of their external similarity the *Gymnotidae* have nothing to do with the eels (*Anguilla*). They resemble the latter in the elongation of the body, the large number of vertebrae (240 in *Gymnotus*), and the absence of pelvic fins; but they differ in all the more important characters of internal structure. They are in fact allied to the carps or *Cyprinidae* and the cat-fishes or *Siluridae*. In common with these two families and the *Characinidae* of Africa and South America, the *Gymnotidae* possess the peculiar structures called *ossicula auditus* or Weberian ossicles. These are a chain of small bones belonging to the first four vertebrae, which are much modified, and connecting the air-bladder with the auditory organs. Such an agreement in the structure of so complicated and specialized an apparatus can only be the result of a community of descent of the families possessing it. Accordingly these families are now placed together in a distinct sub-order, the Ostariophysi. The *Gymnotidae* are strongly modified and degraded *Characinidae*. In them the dorsal and caudal fins are very rudimentary or absent, and the anal is very long, extending from the anus, which is under the head or throat, to the end of the body.

Gymnotus is the only genus of the family which possesses electric organs. These extend the whole length of the tail, which is four-fifths of the body. They are modifications of the lateral muscles and are supplied with numerous branches of the spinal nerves. They consist of longitudinal columns, each composed of an immense number of "electric plates." The posterior end of the organ is positive, the anterior negative, and the current passes from the tail to the head. The maximum shock is given when the head and tail of the *Gymnotus* are in contact with different points in the surface of some other animal. *Gymnotus electricus* attains a length of 3 ft. and the thickness of a man's thigh, and frequents the marshes of Brazil and the Guianas, where it is regarded with terror, owing to the formidable electrical apparatus with which it is provided. When this natural battery is discharged in a favourable position, it is sufficiently powerful to stun the largest animal; and according to A. von Humboldt, it has been found necessary to change the line of certain roads passing through the pools frequented by the electric eels. These fish are eaten by the Indians, who, before attempting to capture them, seek to exhaust their electrical power by driving horses into the ponds. By repeated discharges upon these they gradually expend this marvellous force; after which, being defenceless, they become timid, and approach the edge for shelter, when they fall an easy prey to the harpoon. It is only after long rest and abundance of food that the fish is able to resume the use of its subtle weapon. Humboldt's description of this method of capturing the fish has not, however, been verified by recent travellers.

ELECTRICITY. This article is devoted to a general sketch of the history of the development of electrical knowledge on both the theoretical and the practical sides. The two great branches of electrical theory which concern the phenomena of electricity at rest, or "frictional" or "static" electricity, and of electricity in motion, or electric currents, are treated in two separate articles, Electrostatics and Electrokinetics. The phenomena attendant on the passage of electricity through solids, through liquids and through gases, are described in the article Conduction, Electric, and also Electrolysis, and the propagation of electrical vibrations in Electric Waves. The interconnexion of magnetism (which has an article to itself) and electricity is discussed in Electromagnetism, and these manifestations in nature in Atmospheric Electricity; Aurora Polaris and Magnetism, Terrestrial. The general principles of electrical engineering will be found in Electricity Supply, and further details respecting the generation and use of electrical power are given in such articles as Dynamo; Motors, Electric; Transformers; Accumulator; Power Transmission: *Electric*; Traction; Lighting: *Electric*; Electrochemistry and Electrometallurgy. The principles of telegraphy (land, submarine and wireless) and of telephony are discussed in the articles Telegraph and Telephone, and various electrical instruments are treated in separate articles such as Amperemeter; Electrometer; Galvanometer; Voltmeter; Wheatstone's Bridge; Potentiometer; Meter, Electric; Electrophorus; Leyden Jar; &c. 180

The term "electricity" is applied to denote the physical agency which exhibits itself by effects of attraction and repulsion when particular substances are rubbed or heated, also in certain chemical and physiological actions and in connexion with moving magnets and metallic circuits. The name is derived from the word *electrica*, first used by William Gilbert (1544-1603) in his epoch-making treatise *De magnete, magneticisque corporibus, et de magno magnete tellure*, published in 1600,[1] to denote substances which possess a similar property to amber (= *electrum*, from ἤλεκτρον) of attracting light objects when rubbed. Hence the phenomena came to be collectively called electrical, a term first used by William Barlowe, archdeacon of Salisbury, in 1618, and the study of them, electrical science.

Historical Sketch.

Gilbert was the first to conduct systematic scientific experiments on electrical phenomena. Prior to his date the scanty knowledge possessed by the ancients and enjoyed in the middle ages began and ended with facts said to have been familiar to Thales of Miletus (600 B.C.) and mentioned by Theophrastus (321 B.C.) and Pliny (A.D. 70), namely, that amber, jet and one or two other substances possessed the power, when rubbed, of attracting fragments of straw, leaves or feathers. Starting with careful and accurate observations on facts concerning the mysterious properties of amber and the lodestone, Gilbert laid the foundations of modern electric and magnetic science on the true experimental and inductive basis. The subsequent history of electricity may be divided into four well-marked periods. The first extends from the date of publication of Gilbert's great treatise in 1600 to the invention by Volta of the voltaic pile and the first production of the electric current in 1799. The second dates from Volta's discovery to the discovery by Faraday in 1831 of the induction of electric currents and the creation of currents by the motion of conductors in magnetic fields, which initiated the era of modern electrotechnics. The third covers the period between 1831 and Clerk Maxwell's enunciation of the electromagnetic theory of light in 1865 and the invention of the self-exciting dynamo, which marks another great epoch in the development of the subject; and the fourth comprises the modern development of electric theory and of absolute quantitative measurements, and above all, of the applications of this knowledge in electrical engineering. We shall sketch briefly the historical progress during these various stages, and also the growth of electrical theories of electricity during that time.

First Period.—Gilbert was probably led to study the phenomena of the attraction of iron by the lodestone in consequence of his conversion to the Copernican theory of the earth's motion, and thence proceeded to study the attractions produced by amber. An account of his electrical discoveries is given in the *De magnete*, lib. ii. cap. 2.[2] He invented the *versorium* or electrical needle and proved that innumerable bodies he called *electrica*, when rubbed, can attract the needle of the versorium (see Electroscope). Robert Boyle added many new facts and gave an account of them in his book, *The Origin of Electricity*. He showed that the attraction between the rubbed body and the test object is mutual. Otto von Guericke (1602-1686) constructed the first electrical machine with a revolving ball of sulphur (see Electrical Machine), and noticed that light objects were repelled after being attracted by excited electrics. Sir Isaac Newton substituted a ball of glass for sulphur in the electrical machine and made other not unimportant additions to electrical knowledge. Francis Hawksbee (d. 1713) published in his book *Physico-Mechanical Experiments* (1709), and in several Memoirs in the *Phil. Trans.* about 1707, the results of his electrical inquiries. He showed that light was produced when mercury was shaken up in a glass tube exhausted of its air. Dr Wall observed the spark and crackling sound when warm amber was rubbed, and compared them with thunder and lightning (*Phil. Trans.*, 1708, 26, p. 69). Stephen Gray (1696-1736) noticed in 1720 that electricity could be excited by the friction of hair, silk, wool, paper and other bodies. In 1729 Gray made the important discovery that some bodies were conductors and others non-conductors of electricity. In conjunction with his friend Granville Wheeler (d. 1770), he conveyed the electricity from rubbed glass, a distance of 886 ft., along a string supported on silk threads (*Phil. Trans.*, 1735-1736, 39, pp. 16, 166 and 400). Jean Théophile Desaguliers (1683-1744) announced soon after that electrics were non-conductors, and conductors were non-electrics. C.F. de C. du Fay (1699-1739) made the great discovery that electricity is of two kinds, vitreous and resinous (*Phil. Trans.*, 1733, 38, p. 263), the first being produced when glass, crystal, &c. are rubbed with silk, and the second when resin, amber, silk or paper, &c. are excited by friction with flannel. He also discovered that a body charged with positive or negative electricity repels a body free to move when the latter is charged with electricity of like sign, but attracts it if it is charged with electricity of opposite sign, *i.e.* positive repels positive and negative repels negative, but positive attracts negative. It is to du Fay also that we owe the abolition of the distinction between electrics and non-electrics. He showed that all substances could be electrified by friction, but that to electrify conductors they must be insulated or supported on non-conductors. Various improvements were made in the electrical machine, and thereby experimentalists were provided with the means of generating strong electrification; C.F. Ludolff (1707-1763) of Berlin in 1744 succeeded in igniting ether with the electric spark (*Phil. Trans.*, 1744, 43, p. 167).

For a very full list of the papers and works of these early electrical philosophers, the reader is referred to the bibliography on Electricity in Dr Thomas Young's *Natural Philosophy*, vol. ii. p. 415.

In 1745 the important invention of the Leyden jar or condenser was made by E.G. von Kleist of Kammin, and almost simultaneously by Cunaeus and Pieter van Musschenbroek (1692-1761) of Leiden (see Leyden Jar). Sir William Watson (1715-1787) in England first observed the flash of light when a Leyden jar is discharged, and he and Dr John Bevis (1695-1771) suggested coating the jar inside and outside with tinfoil. Watson carried out elaborate experiments to discover how far the electric discharge of the jar could be conveyed along metallic wires and was able to accomplish it for a distance of 2 m., making the important observation that the electricity appeared to be transmitted instantaneously.

Franklin's Researches.—Benjamin Franklin (1706-1790) was one of the great pioneers of electrical science, and made the ever-memorable experimental identification of lightning and electric spark. He argued that electricity is not created by friction, but merely collected from its state of diffusion through other matter by which it is attracted. He asserted that the glass globe, when rubbed, attracted the electrical fire, and took it from the rubber, the same globe being disposed, when the friction ceases, to give out its electricity to any body which has less. In the case of the charged Leyden jar, he asserted that the inner coating of tinfoil had 181 received more than its ordinary quantity of electricity, and was therefore electrified positively, or plus, while the outer coating of tinfoil having had its ordinary quantity of electricity diminished, was electrified negatively, or minus. Hence the cause of the shock and spark when the jar is discharged, or when the superabundant or plus electricity of the inside is transferred by a conducting body to the defective or minus electricity of the outside. This theory of the Leyden phial Franklin supported very ingeniously by showing that the outside and the inside coating possessed electricities of opposite sign, and that, in charging it, exactly as much electricity is added on one side as is subtracted from the other. The abundant discharge of electricity by points was observed by Franklin is his earliest experiments, and also the power of points to conduct it copiously from an electrified body. Hence he was furnished with a simple method of collecting electricity from other bodies, and he was enabled to perform those remarkable experiments which are chiefly connected with his name. Hawksbee, Wall and J.A. Nollet (1700-1770) had successively suggested the identity of lightning and the electric spark, and of thunder and the snap of the spark. Previously to the year 1750, Franklin drew up a statement, in which he showed that all the general phenomena and effects which were produced by electricity had their counterparts in lightning. After waiting some time for the erection of a spire at Philadelphia, by means of which he hoped to bring down the electricity of a thunderstorm, he conceived the idea of sending up a kite among thunder-clouds. With this view he made a small cross of two small light strips of cedar, the arms being sufficiently long to reach to the four corners of a large thin silk handkerchief when extended. The corners of the handkerchief were tied to the extremities of the cross, and when the body of the kite was thus formed, a tail, loop and string were added to it. The body was made of silk to enable it to bear the violence and wet of a thunderstorm. A very sharp pointed wire was fixed at the top of the upright stick of the cross, so as to rise a foot or more above the wood. A silk ribbon was tied to the end of the twine next the hand, and a key suspended at the junction of the twine and silk. In company with his son, Franklin raised the kite like a common one, in the first thunderstorm, which happened in the month of June 1752. To keep the silk ribbon dry, he stood within a door, taking care that the twine did not touch the frame of the door; and when the thunder-clouds came over the kite he watched the state of the string. A cloud passed without any electrical indications, and he began to despair of success. At last, however, he saw the loose filaments of the twine standing out every way, and he found them to be attracted by the approach of his finger. The suspended key gave a spark on the application of his knuckle, and when the string had become wet with the rain the electricity became abundant. A Leyden jar was charged at the key, and by the electric fire thus obtained spirits were inflamed, and many other experiments performed which had been formerly made by excited electrics. In subsequent trials with another apparatus, he found that the clouds were sometimes positively and sometimes negatively electrified, and so demonstrated the perfect identity of lightning and electricity. Having thus succeeded in drawing the electric fire from the clouds, Franklin conceived the idea of protecting buildings from lightning by erecting on their highest parts pointed iron wires or conductors communicating with the ground. The electricity of a hovering or a passing cloud would thus be carried off slowly and silently; and if the cloud was highly charged, the lightning would strike in preference the elevated conductors.[3] The most important of Franklin's electrical writings are his *Experiments and Observations on Electricity made at Philadelphia*, 1751-1754; his *Letters on Electricity*; and various memoirs and letters in the *Phil. Trans.* from 1756 to 1760.

About the same time that Franklin was making his kite experiment in America, T.F. Dalibard (1703-1779) and others in France had erected a long iron rod at Marli, and obtained results agreeing with those of Franklin. Similar investigations were pursued by many others, among whom Father G.B. Beccaria (1716-1781) deserves especial mention. John Canton (1718-1772) made the important contribution to knowledge that electricity of either sign could be produced on nearly any body by friction with appropriate substances, and that a rod of glass roughened on one half was excited negatively in the rough part and positively in the smooth part by friction with the same rubber. Canton first suggested the use of an amalgam of mercury and tin for use with glass cylinder electrical machines to improve their action. His most important discovery, however, was that of electrostatic induction, the fact that one electrified body can produce charges of electricity upon another insulated body, and that when this last is touched it is left electrified with a charge of opposite sign to that of the inducing charge (*Phil. Trans.*, 1753-1754). We shall make mention lower down of Canton's contributions to electrical theory. Robert Symmer (d. 1763) showed that quite small differences determined the sign of the electrification that was generated by the friction of two bodies one against the other. Thus wearing a black and a white silk stocking one over the other, he found they were electrified oppositely when rubbed and drawn off, and that such a rubbed silk stocking when deposited in a Leyden jar gave up its electrification to the jar (*Phil. Trans.*, 1759). Ebenezer Kinnersley (1711-1778) of Philadelphia made useful observations on the elongation and fusion of iron wires by electrical discharges (*Phil. Trans.*, 1763). A contemporary of Canton and co-discoverer with him of the facts of electrostatic induction was the Swede, Johann Karl Wilcke (1732-1796), then resident in Germany, who in 1762 published an account of experiments in which a metal plate held above the upper surface of a glass table was subjected to the action of a charge on an electrified metal plate held below the glass (*Kon. Schwedische Akad. Abhandl.*, 1762, 24, p. 213).

Pyro-electricity.—The subject of pyro-electricity, or the power possessed by some minerals of becoming electrified when merely heated, and of exhibiting positive and negative electricity, now began to attract notice. It is possible that the *lyncurium* of the ancients, which according to Theophrastus attracted light bodies, was tourmaline, a mineral found in Ceylon, which had been christened by the Dutch with the name of *aschentrikker*, or the attractor of ashes. In 1717 Louis Lémery exhibited to the Paris Academy of Sciences a stone from Ceylon which attracted light bodies; and Linnaeus in mentioning his experiments gives the stone the name of *lapis electricus*. Giovanni Caraffa, duca di Noja (1715-1768), was led in 1758 to purchase some of the stones called tourmaline in Holland, and, assisted by L.J.M. Daubenton and Michel Adanson, he made a series of experiments with them, a description of which he gave in a letter to G.L.L. Buffon in 1759. The subject, however, had already engaged the attention of the German philosopher, F.U.T. Aepinus, who published an account of them in 1756. Hitherto nothing had been said respecting the necessity of heat to excite the tourmaline; but it was shown by Aepinus that a temperature between 99½° and 212° Fahr. was requisite for the development of its attractive powers. Benjamin Wilson (*Phil. Trans.*, 1763, &c.), J. Priestley, and Canton continued the investigation, but it was reserved for the Abbé Haüy to throw a clear light on this curious branch of the science (*Traité de minéralogie*, 1801). He found that the electricity of the tourmaline decreased rapidly from the summits or poles towards the middle of the crystal, where it was imperceptible; and he discovered that if a tourmaline is broken into any number of fragments, each fragment, when excited, has two opposite poles. Haüy discovered the same property in the Siberian and Brazilian topaz, borate of magnesia, mesotype, prehnite, sphene and calamine. He also found that the polarity which minerals receive from heat has a relation to the secondary forms of their crystals—the tourmaline, for example, having its resinous pole at the summit of the crystal which has three faces. In the other pyro-electric crystals above mentioned, Haüy detected the same deviation from the rules of symmetry in their secondary crystals which 182
occurs in tourmaline. C.P. Brard (1788-1838) discovered that pyro-electricity was a property of axinite; and it was afterwards detected in other minerals. In repeating and extending the experiments of Haüy much later, Sir David Brewster discovered that various artificial salts were pyro-electric, and he mentions the tartrates of potash and soda and tartaric acid as exhibiting this property in a very strong degree. He also made many experiments with the tourmaline when cut into thin slices, and reduced to the finest powder, in which state each particle preserved its pyro-electricity; and he showed that scolezite and mesolite, even when deprived of their water of crystallization and reduced to powder, retain their property of becoming electrical by heat.

When this white powder is heated and stirred about by any substance whatever, it collects in masses like new-fallen snow, and adheres to the body with which it is stirred.

For Sir David Brewster’s work on pyro-electricity, see *Trans. Roy. Soc. Edin.*, 1845, also *Phil. Mag.*, Dec. 1847. The reader will also find a full discussion on the subject in the *Treatise on Electricity*, by A. de la Rive, translated by C.V. Walker (London, 1856), vol. ii. part v. ch. i.

Animal electricity.—The observation that certain animals could give shocks resembling the shock of a Leyden jar induced a closer examination of these powers. The ancients were acquainted with the benumbing power of the torpedo-fish, but it was not till 1676 that modern naturalists had their attention again drawn to the fact. E. Bancroft was the first person who distinctly suspected that the effects of the torpedo were electrical. In 1773 John Walsh (d. 1795) and Jan Ingenhousz (1730-1799) proved by many curious experiments that the shock of the torpedo was an electrical one (*Phil. Trans.*, 1773-1775); and John Hunter (id. 1773, 1775) examined and described the anatomical structure of its electrical organs. A. von Humboldt and Gay-Lussac (*Ann. Chim.*, 1805), and Etienne Geoffroy Saint-Hilaire (*Gilb. Ann.*, 1803) pursued the subject with success; and Henry Cavendish (*Phil. Trans.*, 1776) constructed an artificial torpedo, by which he imitated the actions of the living animal. The subject was also investigated (*Phil. Trans.*, 1812, 1817) by Dr T.J. Todd (1789-1840), Sir Humphry Davy (id. 1829), John Davy (id. 1832, 1834, 1841) and Faraday (*Exp. Res.*, vol. ii.). The power of giving electric shocks has been discovered also in the *Gymnotus electricus* (electric eel), the *Malapterurus electricus*, the *Trichiurus electricus*, and the *Tetraodon electricus*. The most interesting and the best known of these singular fishes is the *Gymnotus* or Surinam eel. Humboldt gives a very graphic account of the combats which are carried on in South America between the gymnoti and the wild horses in the vicinity of Calabozo.

Cavendish’s Researches.—The work of Henry Cavendish (1731-1810) entitles him to a high place in the list of electrical investigators. A considerable part of Cavendish’s work was rescued from oblivion in 1879 and placed in an easily accessible form by Professor Clerk Maxwell, who edited the original manuscripts in the possession of the duke of Devonshire.[4] Amongst Cavendish’s important contributions were his exact measurements of electrical capacity. The leading idea which distinguishes his work from that of his predecessors was his use of the phrase “degree of electrification” with a clear scientific definition which shows it to be equivalent in meaning to the modern term “electric potential.” Cavendish compared the capacity of different bodies with those of conducting spheres of known diameter and states these capacities in “globular inches,” a globular inch being the capacity of a sphere 1 in. in diameter. Hence his measurements are all directly comparable with modern electrostatic measurements in which the unit of capacity is that of a sphere 1 centimetre in radius. Cavendish measured the capacity of disks and condensers of various forms, and proved that the capacity of a Leyden pane is proportional to the surface of the tinfoil and inversely as the thickness of the glass. In connexion with this subject he anticipated one of Faraday’s greatest discoveries, namely, the effect of the dielectric or insulator upon the capacity of a condenser formed with it, in other words, made the discovery of specific inductive capacity (see *Electrical Researches*, p. 183). He made many measurements of the electric conductivity of different solids and liquids, by comparing the intensity of the electric shock taken through his body and various conductors. He seems in this way to have educated in himself a very precise “electrical sense,” making use of his own nervous system as a kind of physiological galvanometer. One of the most important investigations he made in this way was to find out, as he expressed it, “what power of the velocity the resistance is proportional to.” Cavendish meant by the term “velocity” what we now call the current, and by “resistance” the electromotive force which maintains the current. By various experiments with liquids in tubes he found this power was nearly unity. This result thus obtained by Cavendish in January 1781, that the current varies in direct proportion to the electromotive force, was really an anticipation of the fundamental law of electric flow, discovered independently by G.S. Ohm in 1827, and since known as Ohm’s Law. Cavendish also enunciated in 1776 all the laws of division of electric current between circuits in parallel, although they are generally supposed to have been first given by Sir C. Wheatstone. Another of his great investigations was the determination of the law according to which electric force varies with the distance. Starting from the fact that if an electrified globe, placed within two hemispheres which fit over it without touching, is brought in contact with these hemispheres, it gives up the whole of its charge to them—in other words, that the charge on an electrified body is wholly on the surface—he was able to deduce by most ingenious reasoning the law that electric force varies inversely as the square of the distance. The accuracy of his measurement, by which he established within 2% the above law, was only limited by the sensibility, or rather insensibility, of the pith ball electrometer, which was his only means of detecting the electric charge.[5] In the accuracy of his quantitative measurements and the range of his researches and his combination of mathematical and physical knowledge, Cavendish may not inaptly be described as the Kelvin of the 18th century. Nothing but his curious indifference to the publication of his work prevented him from securing earlier recognition for it.

Coulomb’s Work.—Contemporary with Cavendish was C.A. Coulomb (1736-1806), who in France addressed himself to the same kind of exact quantitative work as Cavendish in England. Coulomb has made his name for ever famous by his invention and application of his torsion balance to the experimental verification of the fundamental law of electric attraction, in which, however, he was anticipated by Cavendish, namely, that the force of attraction between two small electrified spherical bodies varies as the product of their charges and inversely as the square of the distance of their centres. Coulomb’s work received better publication than Cavendish’s at the time of its accomplishment, and provided a basis on which mathematicians could operate. Accordingly the close of the 18th century drew into the arena of electrical investigation on its mathematical side P.S. Laplace, J.B. Biot, and above all, S.D. Poisson. Adopting the hypothesis of two fluids, Coulomb investigated experimentally and theoretically the distribution of electricity on the surface of bodies by means of his proof plane. He determined the law of distribution between two conducting bodies in contact; and measured with his proof plane the density of the electricity at different points of two spheres in contact, and enunciated an important law. He ascertained the distribution of electricity among several spheres (whether equal or unequal) placed in contact in a straight line; and he measured the distribution of electricity on the surface of a cylinder, and its distribution between a sphere and cylinder of different lengths but of the same diameter. His experiments on the dissipation of electricity possess also a high value. He found that the 183 momentary dissipation was proportional to the degree of electrification at the time, and that, when the charge was moderate, its dissipation was not altered in bodies of different kinds or shapes. The temperature and pressure of the atmosphere did not produce any sensible change; but he concluded that the dissipation was nearly proportional to the cube of the quantity of moisture in the air.[6] In examining the dissipation which takes place along imperfectly insulating substances, he found that a thread of gum-lac was the most perfect of all insulators; that it insulated ten times as well as a dry silk thread; and that a silk thread covered with fine sealing-wax insulated as powerfully as gum-lac when it had four times its length. He found also that the dissipation of electricity along insulators was chiefly owing to adhering moisture, but in some measure also to a slight conducting power. For his memoirs see *Mém. de math. et phys. de l'acad. de sc.*, 1785, &c.

Second Period.—We now enter upon the second period of electrical research inaugurated by the epoch-making discovery of Alessandro Volta (1745-1827). L. Galvani had made in 1790 his historic observations on the muscular contraction produced in the bodies of recently killed frogs when an electrical machine was being worked in the same room, and described them in 1791 (*De viribus electricitatis in motu musculari commentarius*, Bologna, 1791). Volta followed up these observations with rare philosophic insight and experimental skill. He showed that all conductors liquid and solid might be divided into two classes which he called respectively conductors of the first and of the second class, the first embracing metals and carbon in its conducting form, and the second class, water, aqueous solutions of various kinds, and generally those now called electrolytes. In the case of conductors of the first class he proved by the use of the condensing electroscope, aided probably by some form of multiplier or doubler, that a difference of potential (see Electrostatics) was created by the mere contact of two such conductors, one of them being positively electrified and the other negatively. Volta showed, however, that if a series of bodies of the first class, such as disks of various metals, are placed in contact, the potential difference between the first and the last is just the same as if they are immediately in contact. There is no accumulation of potential. If, however, pairs of metallic disks, made, say, of zinc and copper, are alternated with disks of cloth wetted with a conductor of the second class, such, for instance, as dilute acid or any electrolyte, then the effect of the feeble potential difference between one pair of copper and zinc disks is added to that of the potential difference between the next pair, and thus by a sufficiently long series of pairs any required difference of potential can be accumulated.

The Voltaic Pile.—This led him about 1799 to devise his famous voltaic pile consisting of disks of copper and zinc or other metals with wet cloth placed between the pairs. Numerous examples of Volta's original piles at one time existed in Italy, and were collected together for an exhibition held at Como in 1899, but were unfortunately destroyed by a disastrous fire on the 8th of July 1899. Volta's description of his pile was communicated in a letter to Sir Joseph Banks, president of the Royal Society of London, on the 20th of March 1800, and was printed in the *Phil. Trans.*, vol. 90, pt. 1, p. 405. It was then found that when the end plates of Volta's pile were connected to an electroscope the leaves diverged either with positive or negative electricity. Volta also gave his pile another form, the *couronne des tasses* (crown of cups), in which connected strips of copper and zinc were used to bridge between cups of water or dilute acid. Volta then proved that all metals could be arranged in an electromotive series such that each became positive when placed in contact with the one next below it in the series. The origin of the electromotive force in the pile has been much discussed, and Volta's discoveries gave rise to one of the historic controversies of science. Volta maintained that the mere contact of metals was sufficient to produce the electrical difference of the end plates of the pile. The discovery that chemical action was involved in the process led to the advancement of the chemical theory of the pile and this was strengthened by the growing insight into the principle of the conservation of energy. In 1851 Lord Kelvin (Sir W. Thomson), by the use of his then newly-invented electrometer, was able to confirm Volta's observations on contact electricity by irrefutable evidence, but the contact theory of the voltaic pile was then placed on a basis consistent with the principle of the conservation of energy. A.A. de la Rive and Faraday were ardent supporters of the chemical theory of the pile, and even at the present time opinions of physicists can hardly be said to be in entire accordance as to the source of the electromotive force in a voltaic couple or pile.[7]

Improvements in the form of the voltaic pile were almost immediately made by W. Cruickshank (1745-1800), Dr W.H. Wollaston and Sir H. Davy, and these, together with other eminent continental chemists, such as A.F. de Fourcroy, L.J. Thénard and J.W. Ritter (1776-1810), ardently prosecuted research with the new instrument. One of the first discoveries made with it was its power to electrolyse or chemically decompose certain solutions. William Nicholson (1753-1815) and Sir Anthony Carlisle (1768-1840) in 1800 constructed a pile of silver and zinc plates, and placing the terminal wires in water noticed the evolution from these wires of bubbles of gas, which they proved to be oxygen and hydrogen. These two gases, as Cavendish and James Watt had shown in 1784, were actually the constituents of water. From that date it was clearly recognized that a fresh implement of great power had been given to the chemist. Large voltaic piles were then constructed by Andrew Crosse (1784-1855) and Sir H. Davy, and improvements initiated by Wollaston and Robert Hare (1781-1858) of Philadelphia. In 1806 Davy communicated to the Royal Society of London a celebrated paper on some "Chemical Agencies of Electricity," and after providing himself at the Royal Institution of London with a battery of several hundred cells, he announced in 1807 his great discovery of the electrolytic decomposition of the alkalis, potash and soda, obtaining therefrom the metals potassium and sodium. In July 1808 Davy laid a request before the managers of the Royal Institution that they would set on foot a subscription for the purchase of a specially large voltaic battery; as a result he was provided with one of 2000 pairs of plates, and the first experiment performed with it was the production of the electric arc light between carbon poles. Davy followed up his initial work with a long and brilliant series of electrochemical investigations described for the most part in the *Phil. Trans.* of the Royal Society.

Magnetic Action of Electric Current.—Noticing an analogy between the polarity of the voltaic pile and that of the magnet, philosophers had long been anxious to discover a relation between the two, but twenty years elapsed after the invention of the pile before Hans Christian Oersted (1777-1851), professor of natural philosophy in the university of Copenhagen, made in 1819 the discovery which has immortalized his name. In the *Annals of Philosophy* (1820, 16, p. 273) is to be found an English translation of Oersted's original Latin essay (entitled "Experiments on the Effect of a Current of Electricity on the Magnetic Needle"), dated the

21st of July 1820, describing his discovery. In it Oersted describes the action he considers is taking place around the conductor joining the extremities of the pile; he speaks of it as the electric conflict, and says: "It is sufficiently evident that the electric conflict is not confined to the conductor, but is dispersed pretty widely in the circumjacent space. We may likewise conclude that this conflict performs circles round the wire, for without this condition it seems impossible that one part of the wire when placed below the magnetic needle should drive its pole to the east, and when placed above it, to the west." Oersted's important discovery was the fact that when a wire joining the end plates of a voltaic pile is held near a pivoted magnet or compass needle, the latter is deflected and places itself more or less transversely to the wire, the direction depending upon whether the wire is above or below the needle, and on the manner in which the copper or zinc ends of the pile are connected to it. It is clear, moreover, that Oersted clearly recognized the existence of what is now called the magnetic field round the conductor. This discovery of Oersted, like that of Volta, stimulated philosophical investigation in a high degree.

Electrodynamics.—On the 2nd of October 1820, A.M. Ampère presented to the French Academy of Sciences an important memoir,[8] in which he summed up the results of his own and D.F.J. Arago's previous investigations in the new science of electromagnetism, and crowned that labour by the announcement of his great discovery of the dynamical action between conductors conveying the electric currents. Ampère in this paper gave an account of his discovery that conductors conveying electric currents exercise a mutual attraction or repulsion on one another, currents flowing in the same direction in parallel conductors attracting, and those in opposite directions repelling. Respecting this achievement when developed in its experimental and mathematical completeness, Clerk Maxwell says that it was "perfect in form and unassailable in accuracy." By a series of well-chosen experiments Ampère established the laws of this mutual action, and not only explained observed facts by a brilliant train of mathematical analysis, but predicted others subsequently experimentally realized. These investigations led him to the announcement of the fundamental law of action between elements of current, or currents in infinitely short lengths of linear conductors, upon one another at a distance; summed up in compact expression this law states that the action is proportional to the product of the current strengths of the two elements, and the lengths of the two elements, and inversely proportional to the square of the distance between the two elements, and also directly proportional to a function of the angles which the line joining the elements makes with the directions of the two elements respectively. Nothing is more remarkable in the history of discovery than the manner in which Ampère seized upon the right clue which enabled him to disentangle the complicated phenomena of electrodynamics and to deduce them all as a consequence of one simple fundamental law, which occupies in electrodynamics the position of the Newtonian law of gravitation in physical astronomy.

In 1821 Michael Faraday (1791-1867), who was destined later on to do so much for the science of electricity, discovered electromagnetic rotation, having succeeded in causing a wire conveying a voltaic current to rotate continuously round the pole of a permanent magnet.[9] This experiment was repeated in a variety of forms by A.A. De la Rive, Peter Barlow (1776-1862), William Ritchie (1790-1837), William Sturgeon (1783-1850), and others; and Davy (*Phil. Trans.*, 1823) showed that when two wires connected with the pole of a battery were dipped into a cup of mercury placed on the pole of a powerful magnet, the fluid rotated in opposite directions about the two electrodes.

Electromagnetism.—In 1820 Arago (*Ann. Chim. Phys.*, 1820, 15, p. 94) and Davy (*Annals of Philosophy*, 1821) discovered independently the power of the electric current to magnetize iron and steel. Félix Savary (1797-1841) made some very curious observations in 1827 on the magnetization of steel needles placed at different distances from a wire conveying the discharge of a Leyden jar (*Ann. Chim. Phys.*, 1827, 34). W. Sturgeon in 1824 wound a copper wire round a bar of iron bent in the shape of a horseshoe, and passing a voltaic current through the wire showed that the iron became powerfully magnetized as long as the connexion with the pile was maintained (*Trans. Soc. Arts*, 1825). These researches gave us the electromagnet, almost as potent an instrument of research and invention as the pile itself (see Electromagnetism).

Ampère had already previously shown that a spiral conductor or solenoid when traversed by an electric current possesses magnetic polarity, and that two such solenoids act upon one another when traversed by electric currents as if they were magnets. Joseph Henry, in the United States, first suggested the construction of what were then called intensity electromagnets, by winding upon a horseshoe-shaped piece of soft iron many superimposed windings of copper wire, insulated by covering it with silk or cotton, and then sending through the coils the current from a voltaic battery. The dependence of the intensity of magnetization on the strength of the current was subsequently investigated (*Pogg. Ann. Phys.*, 1839, 47) by H.F.E. Lenz (1804-1865) and M.H. von Jacobi (1801-1874). J.P. Joule found that magnetization did not increase proportionately with the current, but reached a maximum (*Sturgeon's Annals of Electricity*, 1839, 4). Further investigations on this subject were carried on subsequently by W.E. Weber (1804-1891), J.H.J. Müller (1809-1875), C.J. Dub (1817-1873), G.H. Wiedemann (1826-1899), and others, and in modern times by H.A. Rowland (1848-1901), Shelford Bidwell (b. 1848), John Hopkinson (1849-1898), J.A. Ewing (b. 1855) and many others. Electric magnets of great power were soon constructed in this manner by Sturgeon, Joule, Henry, Faraday and Brewster. Oersted's discovery in 1819 was indeed epoch-making in the degree to which it stimulated other research. It led at once to the construction of the galvanometer as a means of detecting and measuring the electric current in a conductor. In 1820 J.S.C. Schweigger (1779-1857) with his "multiplier" made an advance upon Oersted's discovery, by winding the wire conveying the electric current many times round the pivoted magnetic needle and thus increasing the deflection; and L. Nobili (1784-1835) in 1825 conceived the ingenious idea of neutralizing the directive effect of the earth's magnetism by employing a pair of magnetized steel needles fixed to one axis, but with their magnetic poles pointing in opposite directions. Hence followed the astatic multiplying galvanometer.

Electrodynamic Rotation.—The study of the relation between the magnet and the circuit conveying an electric current then led Arago to the discovery of the "magnetism of rotation." He found that a vibrating magnetic compass needle came to rest sooner when placed over a plate of copper than otherwise, and also that a plate of copper rotating under a suspended magnet tended to drag the magnet in the same direction.

The matter was investigated by Charles Babbage, Sir J.F.W. Herschel, Peter Barlow and others, but did not receive a final explanation until after the discovery of electromagnetic induction by Faraday in 1831. Ampère's investigations had led electricians to see that the force acting upon a magnetic pole due to a current in a neighbouring conductor was such as to tend to cause the pole to travel round the conductor. Much ingenuity had, however, to be expended before a method was found of exhibiting such a rotation. Faraday first succeeded by the simple but ingenious device of using a light magnetic needle tethered flexibly to the bottom of a cup containing mercury so that one pole of the magnet was just above the surface of the mercury. On bringing down on to the mercury surface a wire conveying an electric current, and allowing the current to pass through the mercury and out at the bottom, the magnetic pole at once began to rotate round the wire (*Exper. Res.*, 1822, 2, p. 148). Faraday and others then discovered, as already mentioned, means to make the conductor conveying the current rotate round a magnetic pole, and Ampère showed that a magnet 185
could be made to rotate on its own axis when a current was passed through it. The difficulty in this case consisted in discovering means by which the current could be passed through one half of the magnet without passing it through the other half. This, however, was overcome by sending the current out at the centre of the magnet by means of a short length of wire dipping into an annular groove containing mercury. Barlow, Sturgeon and others then showed that a copper disk could be made to rotate between the poles of a horseshoe magnet when a current was passed through the disk from the centre to the circumference, the disk being rendered at the same time freely movable by making a contact with the circumference by means of a mercury trough. These experiments furnished the first elementary forms of electric motor, since it was then seen that rotatory motion could be produced in masses of metal by the mutual action of conductors conveying electric current and magnetic fields. By his discovery of thermo-electricity in 1822 (*Pogg. Ann. Phys.*, 6), T.J. Seebeck (1770-1831) opened up a new region of research (see THERMO-ELECTRICITY). James Cumming (1777-1861) in 1823 (*Annals of Philosophy*, 1823) found that the thermo-electric series varied with the temperature, and J.C.A. Peltier (1785-1845) in 1834 discovered that a current passed across the junction of two metals either generated or absorbed heat.

Ohm's Law.—In 1827 Dr G.S. Ohm (1787-1854) rendered a great service to electrical science by his mathematical investigation of the voltaic circuit, and publication of his paper, *Die galvanische Kette mathematisch bearbeitet*. Before his time, ideas on the measurable quantities with which we are concerned in an electric circuit were extremely vague. Ohm introduced the clear idea of current strength as an effect produced by electromotive force acting as a cause in a circuit having resistance as its quality, and showed that the current was directly proportional to the electromotive force and inversely as the resistance. Ohm's law, as it is called, was based upon an analogy with the flow of heat in a circuit, discussed by Fourier. Ohm introduced the definite conception of the distribution along the circuit of "electroscopic force" or tension (*Spannung*), corresponding to the modern term potential. Ohm verified his law by the aid of thermo-electric piles as sources of electromotive force, and Davy, C.S.M. Pouillet (1791-1868), A.C. Becquerel (1788-1878), G.T. Fechner (1801-1887), R.H.A. Kohlrausch (1809-1858) and others laboured at its confirmation. In more recent times, 1876, it was rigorously tested by G. Chrystal (b. 1851) at Clerk Maxwell's instigation (see *Brit. Assoc. Report*, 1876, p. 36), and although at its original enunciation its meaning was not at first fully apprehended, it soon took its place as the expression of the fundamental law of electrokinetics.

Induction of Electric Currents.—In 1831 Faraday began the investigations on electromagnetic induction which proved more fertile in far-reaching practical consequences than any of those which even his genius gave to the world. These advances all centre round his supreme discovery of the induction of electric currents. Fully familiar with the fact that an electric charge upon one conductor could produce a charge of opposite sign upon a neighbouring conductor, Faraday asked himself whether an electric current passing through a conductor could not in any like manner induce an electric current in some neighbouring conductor. His first experiments on this subject were made in the month of November 1825, but it was not until the 29th of August 1831 that he attained success. On that date he had provided himself with an iron ring, over which he had wound two coils of insulated copper wire. One of these coils was connected with the voltaic battery and the other with the galvanometer. He found that at the moment the current in the battery circuit was started or stopped, transitory currents appeared in the galvanometer circuit in opposite directions. In ten days of brilliant investigation, guided by clear insight from the very first into the meaning of the phenomena concerned, he established experimentally the fact that a current may be induced in a conducting circuit simply by the variation in a magnetic field, the lines of force of which are linked with that circuit. The whole of Faraday's investigations on this subject can be summed up in the single statement that if a conducting circuit is placed in a magnetic field, and if either by variation of the field or by movement or variation of the form of the circuit the total magnetic flux linked with the circuit is varied, an electromotive force is set up in that circuit which at any instant is measured by the rate at which the total flux linked with the circuit is changing.

Amongst the memorable achievements of the ten days which Faraday devoted to this investigation was the discovery that a current could be induced in a conducting wire simply by moving it in the neighbourhood of a magnet. One form which this experiment took was that of rotating a copper disk between the poles of a powerful electric magnet. He then found that a conductor, the ends of which were connected respectively with the centre and edge of the disk, was traversed by an electric current. This important fact laid the foundation for all subsequent inventions which finally led to the production of electromagnetic or dynamo-electric machines.

THIRD PERIOD.—With this supremely important discovery of Faraday's we enter upon the third period of electrical research, in which that philosopher himself was the leading figure. He not only collected the facts concerning electromagnetic induction so industriously that nothing of importance remained for future discovery, and embraced them all in one law of exquisite simplicity, but he introduced his famous conception of lines of force which changed entirely the mode of regarding electrical phenomena. The French mathematicians, Coulomb, Biot, Poisson and Ampère, had been content to accept the fact that electric charges or currents in conductors could exert forces on other charges or conductors at a distance without inquiring into the means by which this action at a distance was produced. Faraday's mind, however, revolted against this notion; he felt intuitively that these distance actions must be the result of unseen operations in the interposed medium. Accordingly when he sprinkled iron filings on a card held over a magnet and revealed the curvilinear system of lines of force (see Magnetism), he regarded these fragments of iron as simple indicators of a physical state in the space already in existence round the magnet. To him a magnet was not simply a bar of steel; it was the core and origin of a system of lines of magnetic force attached to it and moving with it. Similarly he came to see an electrified body as a centre of a system of lines of electrostatic force. All the space round magnets, currents and electric charges was therefore to Faraday the seat of corresponding lines of magnetic or electric force. He proved by systematic experiments that the electromotive forces set up in conductors by their motions in magnetic fields or by the induction of other currents in the field were due to the secondary conductor *cutting* lines of magnetic force. He invented the term "electrotonic state" to signify the total magnetic flux due to a conductor conveying a current, which was linked with any secondary circuit in the field or even with itself.

Faraday's Researches.—Space compels us to limit our account of the scientific work done by Faraday in the succeeding twenty years, in elucidating electrical phenomena and adding to the knowledge thereon, to the very briefest mention. We must refer the reader for further information to his monumental work entitled *Experimental Researches on Electricity*, in three volumes, reprinted from the *Phil. Trans.* between 1831 and 1851. Faraday divided these researches into various series. The 1st and 2nd concern the discovery of magneto-electric induction already mentioned. The 3rd series (1833) he devoted to discussion of the identity of electricity derived from various sources, frictional, voltaic, animal and thermal, and he proved by rigorous experiments the identity and similarity in properties of the electricity generated by these various methods. The 5th series (1833) is occupied with his electrochemical researches. In the 7th series (1834) he defines a number of new terms, such as electrolyte, electrolysis, anode and cathode, &c., in connexion with electrolytic phenomena, which were immediately adopted into the vocabulary of science. His most important contribution at this date was the invention of the voltameter and his enunciation of the laws of electrolysis. 186
The voltameter provided a means of measuring quantity of electricity, and in the hands of Faraday and his successors became an appliance of fundamental importance. The 8th series is occupied with a discussion of the theory of the voltaic pile, in which Faraday accumulates evidence to prove that the source of the energy of the pile must be chemical. He returns also to this subject in the 16th series. In the 9th series (1834) he announced the discovery of the important property of electric conductors, since called their self-induction or inductance, a discovery in which, however, he was anticipated by Joseph Henry in the United States. The 11th series (1837) deals with electrostatic induction and the statement of the important fact of the specific inductive capacity of insulators or dielectrics. This discovery was made in November 1837 when Faraday had no knowledge of Cavendish's previous researches into this matter. The 19th series (1845) contains an account of his brilliant discovery of the rotation of the plane of polarized light by transparent dielectrics placed in a magnetic field, a relation which established for the first time a practical connexion between the phenomena of electricity and light. The 20th series (1845) contains an account of his researches on the universal action of magnetism and diamagnetic bodies. The 22nd series (1848) is occupied with the discussion of magneto-crystallic force and the abnormal behaviour of various crystals in a magnetic field. In the 25th series (1850) he made known his discovery of the magnetic character of oxygen gas, and the important principle that the terms paramagnetic and diamagnetic are relative. In the 26th series (1850) he returned to a discussion of magnetic lines of force, and illuminated the whole subject of the magnetic circuit by his transcendent insight into the intricate phenomena concerned. In 1855 he brought these researches to a conclusion by a general article on magnetic philosophy, having placed the whole subject of magnetism and electromagnetism on an entirely novel and solid basis. In addition to this he provided the means for studying the phenomena not only qualitatively, but also quantitatively, by the profoundly ingenious instruments he invented for that purpose.

Electrical Measurement.—Faraday's ideas thus pressed upon electricians the necessity for the quantitative measurement of electrical phenomena.[10] It has been already mentioned that Schweigger invented in 1820 the "multiplier," and Nobili in 1825 the astatic galvanometer. C.S.M. Pouillet in 1837 contributed the sine and tangent compass, and W.E. Weber effected great improvements in them and in the construction and use of galvanometers. In 1849 H. von Helmholtz devised a tangent galvanometer with two coils. The measurement of electric resistance then engaged the attention of electricians. By his Memoirs in the *Phil. Trans.* in 1843, Sir Charles Wheatstone gave a great impulse to this study. He invented the rheostat and improved the resistance balance, invented by S.H. Christie (1784-1865) in 1833, and subsequently called the Wheatstone Bridge. (See his *Scientific Papers*, published by the Physical Society of London, p. 129.) Weber about this date invented the electrodynamometer, and applied the mirror and scale method of reading deflections, and in co-operation with C.F. Gauss introduced a system of absolute measurement of electric and magnetic phenomena. In 1846 Weber proceeded with improved apparatus to test Ampère's laws of electrodynamics. In 1845 H.G. Grassmann (1809-1877) published (*Pogg. Ann.* vol. 64) his "Neue Theorie der Electrodynamik," in which he gave an elementary law differing from that of Ampère but leading to the same results for closed circuits. In the same year F.E. Neumann published another law. In 1846 Weber announced his famous hypothesis concerning the connexion of electrostatic and electrodynamic phenomena. The work of Neumann and Weber had been stimulated by that of H.F.E. Lenz (1804-1865), whose researches (*Pogg. Ann.*, 1834, 31; 1835, 34) among other results led him to the statement of the law by means of which the direction of the induced current can be predicted from the theory of Ampère, the rule being that the direction of the induced current is always such that its electrodynamic action tends to oppose the motion which produces it.

Neumann in 1845 did for electromagnetic induction what Ampère did for electrodynamics, basing his researches upon the experimental laws of Lenz. He discovered a function, which has been called the potential of one circuit on another, from which he deduced a theory of induction completely in accordance with experiment. Weber at the same time deduced the mathematical laws of induction from his elementary law of electrical action, and with his improved instruments arrived at accurate verifications of the law of induction, which by this time had been developed mathematically by Neumann and himself. In 1849 G.R. Kirchhoff determined experimentally in a certain case the absolute value of the current induced by one circuit in another, and in the same year Erik Edland (1819-1888) made a series of careful experiments on the induction of electric currents which further established received theories. These labours laid the foundation on which was subsequently erected a complete system for the absolute measurement of electric and magnetic quantities, referring them all to the fundamental units of mass, length and time. Helmholtz gave at the same time a mathematical theory of induced currents and a valuable series of experiments in support of them (*Pogg. Ann.*, 1851). This great investigator and luminous expositor just before that time had published his celebrated essay, *Die Erhaltung der Kraft* (“The Conservation of Energy”), which brought to a focus ideas which had been accumulating in consequence of the work of J.P. Joule, J.R. von Mayer and others, on the transformation of various forms of physical energy, and in particular the mechanical equivalent of heat. Helmholtz brought to bear upon the subject not only the most profound mathematical attainments, but immense experimental skill, and his work in connexion with this subject is classical.

Lord Kelvin’s Work.—About 1842 Lord Kelvin (then William Thomson) began that long career of theoretical and practical discovery and invention in electrical science which revolutionized every department of pure and applied electricity. His early contributions to electrostatics and electrometry are to be found described in his *Reprint of Papers on Electrostatics and Magnetism* (1872), and his later work in his collected *Mathematical and Physical Papers*. By his studies in electrostatics, his elegant method of electrical images, his development of the theory of potential and application of the principle of conservation of energy, as well as by his inventions in connexion with electrometry, he laid the foundations of our modern knowledge of electrostatics. His work on the electrodynamic qualities of metals, thermo-electricity, and his contributions to galvanometry, were not less massive and profound. From 1842 onwards to the end of the 19th century, he was one of the great master workers in the field of electrical discovery and research.[11] In 1853 he published a paper “On Transient Electric Currents” (*Phil. Mag.*, 1853 [4], 5, p. 393), in which he applied the principle of the conservation of energy to the discharge of a Leyden jar. He added definiteness to the idea of the self-induction or inductance of an electric circuit, and gave a mathematical expression for the current flowing out of a Leyden jar during its discharge. He confirmed an opinion already previously expressed by Helmholtz and by Henry, that in some circumstances this discharge is oscillatory in nature, consisting of an alternating electric current of high frequency. These theoretical predictions were confirmed and others, subsequently, by the work of B.W. Feddersen (b. 1832), C.A. Paalzow (b. 1823), and it was then seen that the familiar phenomena of the discharge of a Leyden jar provided the means of generating electric oscillations of very 187
high frequency.

Telegraphy.—Turning to practical applications of electricity, we may note that electric telegraphy took its rise in 1820, beginning with a suggestion of Ampère immediately after Oersted’s discovery. It was established by the work of Weber and Gauss at Göttingen in 1836, and that of C.A. Steinheil (1801-1870) of Munich, Sir W.F. Cooke (1806-1879) and Sir C. Wheatstone in England, Joseph Henry and S.F.B. Morse (1791-1872) in the United States in 1837. In 1845 submarine telegraphy was inaugurated by the laying of an insulated conductor across the English Channel by the brothers Brett, and their temporary success was followed by the laying in 1851 of a permanent Dover-Calais cable by T.R. Crampton. In 1856 the project for an Atlantic submarine cable took shape and the Atlantic Telegraph Company was formed with a capital of £350,000, with Sir Charles Bright as engineer-in-chief and E.O.W. Whitehouse as electrician. The phenomena connected with the propagation of electric signals by underground insulated wires had already engaged the attention of Faraday in 1854, who pointed out the Leyden-jar-like action of an insulated subterranean wire. Scientific and practical questions connected with the possibility of laying an Atlantic submarine cable then began to be discussed, and Lord Kelvin was foremost in developing true scientific knowledge on this subject, and in the invention of appliances for utilizing it. One of his earliest and most useful contributions (in 1858) was the invention of the mirror galvanometer. Abandoning the long and somewhat heavy magnetic needles that had been used up to that date in galvanometers, he attached to the back of a very small mirror made of microscopic glass a fragment of magnetized watch-spring, and suspended the mirror and needle by means of a cocoon fibre in the centre of a coil of insulated wire. By this simple device he provided a means of measuring small electric currents far in advance of anything yet accomplished, and this instrument proved not only most useful in pure scientific researches, but at the same time was of the utmost value in connexion with submarine telegraphy. The history of the initial failures and final success in laying the Atlantic cable has been well told by Mr. Charles Bright (see *The Story of the Atlantic Cable*, London, 1903).[12] The first cable laid in 1857 broke on the 11th of August during laying. The second attempt in 1858 was successful, but the cable completed on the 5th of August 1858 broke down on the 20th of October 1858, after 732 messages had passed through it. The third cable laid in 1865 was lost on the 2nd of August 1865, but in 1866 a final success was attained and the 1865 cable also recovered and completed. Lord Kelvin’s mirror galvanometer was first used in receiving signals through the short-lived 1858 cable. In 1867 he invented his beautiful siphon-recorder for receiving and recording the signals through long cables. Later, in conjunction with Prof. Fleeming Jenkin, he devised his automatic curb sender, an appliance for sending signals by means of punched telegraphic paper tape. Lord Kelvin’s contributions to the science of exact electric measurement[13] were enormous. His ampere-balances, voltmeters and electrometers, and double bridge, are elsewhere described in detail (see Amperemeter; Electrometer, and Wheatstone’s Bridge).

Dynamo.—The work of Faraday from 1831 to 1851 stimulated and originated an immense mass of scientific research, but at the same time practical inventors had not been slow to perceive that it was capable of purely technical application. Faraday’s copper disk rotated between the poles of a magnet, and producing thereby an electric current, became the parent of innumerable machines in which mechanical energy was directly converted into the energy of electric currents. Of these machines, originally called magneto-electric machines, one of the first was devised in 1832 by H. Pixii. It consisted of a fixed horseshoe armature wound over with insulated copper wire in front of which revolved about a vertical axis a horseshoe magnet. Pixii, who invented the split tube commutator for converting the alternating current so produced into a continuous current in the external circuit, was followed by J. Saxton, E.M. Clarke, and many others in the development of the above-described magneto-electric machine. In 1857 E.W. Siemens effected a great improvement by inventing a shuttle armature and improving the shape of the field magnet. Subsequently similar machines with electromagnets were introduced by Henry Wilde (b. 1833), Siemens, Wheatstone, W. Ladd and others, and the principle of self-excitation was suggested by Wilde, C.F. Varley (1828-1883), Siemens and Wheatstone (see Dynamo). These machines about 1866 and 1867 began to be constructed on a commercial scale and were employed in the production of the electric light. The discovery of electric-current induction also led to the production of the induction coil (*q.v.*), improved and brought to its present perfection by W. Sturgeon, E.R. Ritchie, N.J. Callan, H.D. Rühmkorff (1803-1877), A.H.L. Fizeau, and more recently by A. Apps and modern inventors. About the same time Fizeau and J.B.L. Foucault devoted attention to the invention of automatic apparatus for the production of Davy's electric arc (see Lighting: *Electric*), and these appliances in conjunction with magneto-electric machines were soon employed in lighthouse work. With the advent of large magneto-electric machines the era of electrotechnics was fairly entered, and this period, which may be said to terminate about 1867 to 1869, was consummated by the theoretical work of Clerk Maxwell.

Maxwell's Researches.—James Clerk Maxwell (1831-1879) entered on his electrical studies with a desire to ascertain if the ideas of Faraday, so different from those of Poisson and the French mathematicians, could be made the foundation of a mathematical method and brought under the power of analysis.[14] Maxwell started with the conception that all electric and magnetic phenomena are due to effects taking place in the dielectric or in the ether if the space be vacuous. The phenomena of light had compelled physicists to postulate a space-filling medium, to which the name ether had been given, and Henry and Faraday had long previously suggested the idea of an electromagnetic medium. The vibrations of this medium constitute the agency called light. Maxwell saw that it was unphilosophical to assume a multiplicity of ethers or media until it had been proved that one would not fulfil all the requirements. He formulated the conception, therefore, of electric charge as consisting in a displacement taking place in the dielectric or electromagnetic medium (see Electrostatics). Maxwell never committed himself to a precise definition of the physical nature of electric displacement, but considered it as defining that which Faraday had called the polarization in the insulator, or, what is equivalent, the number of lines of electrostatic force passing normally through a unit of area in the dielectric. A second fundamental conception of Maxwell was that the electric displacement whilst it is changing is in effect an electric current, and creates, therefore, magnetic force. The total current at any point in a dielectric must be considered as made up of two parts: first, the true conduction current, if it exists; and second, the rate of change of dielectric displacement. The fundamental fact connecting electric currents and magnetic fields is that the line integral of magnetic force taken once round a conductor conveying an electric current is equal to 4 π-times the surface integral of the current density, or to 4 π-times the total current flowing through the closed line round which the integral is taken (see Electrokinetics). A second relation connecting magnetic and electric force is based upon Faraday's fundamental law of 188
induction, that the rate of change of the total magnetic flux linked with a conductor is a measure of the electromotive force created in it (see Electrokinetics). Maxwell also introduced in this connexion the notion of the vector potential. Coupling together these ideas he was finally enabled to prove that the propagation of electric and magnetic force takes place through space with a certain velocity determined by the dielectric constant and the magnetic permeability of the medium. To take a simple instance, if we consider an electric current as flowing in a conductor it is, as Oersted discovered, surrounded by closed lines of magnetic force. If we imagine the current in the conductor to be instantaneously reversed in direction, the magnetic force surrounding it would not be instantly reversed everywhere in direction, but the reversal would be propagated outwards through space with a certain velocity which Maxwell showed was inversely as the square root of the product of the magnetic permeability and the dielectric constant or specific inductive capacity of the medium.

These great results were announced by him for the first time in a paper presented in 1864 to the Royal Society of London and printed in the *Phil. Trans.* for 1865, entitled "A Dynamical Theory of the Electromagnetic Field." Maxwell showed in this paper that the velocity of propagation of an electromagnetic impulse through space could also be determined by certain experimental methods which consisted in measuring the same electric quantity, capacity, resistance or potential in two ways. W.E. Weber had already laid the foundations of the absolute system of electric and magnetic measurement, and proved that a quantity of electricity could be measured either by the force it exercises upon another static or stationary quantity of electricity, or magnetically by the force this quantity of electricity exercises upon a magnetic pole when flowing through a neighbouring conductor. The two systems of measurement were called respectively the electrostatic and the electromagnetic systems (see Units, Physical). Maxwell suggested new methods for the determination of this ratio of the electrostatic to the electromagnetic units, and by experiments of great ingenuity was able to show that this ratio, which is also that of the velocity of the propagation of an electromagnetic impulse through space, is identical with that of light. This great fact once ascertained, it became clear that the notion that electric phenomena are affections of the luminiferous ether was no longer a mere speculation but a scientific theory capable of verification. An immediate deduction from Maxwell's theory was that in transparent dielectrics, the dielectric constant or specific inductive capacity should be numerically equal to the square of the refractive index for very long electric waves. At the time when Maxwell developed his theory the dielectric constants of only a few transparent insulators were known and these were for the most part measured with steady or unidirectional electromotive force. The only refractive indices which had been measured were the optical refractive indices of a number of transparent substances. Maxwell made a comparison between the optical refractive index and the dielectric constant of paraffin wax, and the approximation between the numerical values of the square of the first and that of the last was sufficient to show that there was a basis for further work. Maxwell's electric and magnetic ideas were gathered together in a great mathematical treatise on electricity and magnetism which was published in 1873.[15] This book stimulated in a most remarkable degree theoretical and practical research into the phenomena of electricity and magnetism. Experimental methods were devised for the further exact measurements of the electromagnetic velocity and numerous determinations of the dielectric constants of various solids, liquids and gases, and comparisons of these with the corresponding optical refractive indices were conducted. This early work indicated that whilst there were a number of cases in which the square of optical refractive index for long waves and the dielectric constant of the same substance were sufficiently close to afford an apparent confirmation of Maxwell's theory, yet in other cases there were considerable divergencies. L. Boltzmann (1844-1907) made a large number of determinations for solids and for gases, and the dielectric constants of many solid and liquid substances were determined by N.N. Schiller (b. 1848), P.A. Silow (b. 1850), J. Hopkinson and others. The accumulating determinations of the numerical value of the electromagnetic velocity (v) from the earliest made by Lord Kelvin (Sir W. Thomson) with the aid of King and

McKichan, or those of Clerk Maxwell, W.E. Ayrton and J. Perry, to more recent ones by J.J. Thomson, F. Himstedt, H.A. Rowland, E.B. Rosa, J.S.H. Pellat and H.A. Abraham, showed it to be very close to the best determinations of the velocity of light (see Units, Physical). On the other hand, the divergence in some cases between the square of the optical refractive index and the dielectric constant was very marked. Hence although Maxwell's theory of electrical action when first propounded found many adherents in Great Britain, it did not so much dominate opinion on the continent of Europe.

Fourth Period.—With the publication of Clerk Maxwell's treatise in 1873, we enter fully upon the fourth and modern period of electrical research. On the technical side the invention of a new form of armature for dynamo electric machines by Z.T. Gramme (1826-1901) inaugurated a departure from which we may date modern electrical engineering. It will be convenient to deal with technical development first.

Technical Development.—As far back as 1841 large magneto-electric machines driven by steam power had been constructed, and in 1856 F.H. Holmes had made a magneto machine with multiple permanent magnets which was installed in 1862 in Dungeness lighthouse. Further progress was made in 1867 when H. Wilde introduced the use of electromagnets for the field magnets. In 1860 Dr Antonio Pacinotti invented what is now called the toothed ring winding for armatures and described it in an Italian journal, but it attracted little notice until reinvented in 1870 by Gramme. In this new form of bobbin, the armature consisted of a ring of iron wire wound over with an endless coil of wire and connected to a commutator consisting of copper bars insulated from one another. Gramme dynamos were then soon made on the self-exciting principle. In 1873 at Vienna the fact was discovered that a dynamo machine of the Gramme type could also act as an electric motor and was set in rotation when a current was passed into it from another similar machine. Henceforth the electric transmission of power came within the possibilities of engineering.

Electric Lighting.—In 1876, Paul Jablochkov (1847-1894), a Russian officer, passing through Paris, invented his famous electric candle, consisting of two rods of carbon placed side by side and separated from one another by an insulating material. This invention in conjunction with an alternating current dynamo provided a new and simple form of electric arc lighting. Two years afterwards C.F. Brush, in the United States, produced another efficient form of dynamo and electric arc lamp suitable for working in series (see Lighting: *Electric*), and these inventions of Brush and Jablochkov inaugurated commercial arc lighting. The so-called subdivision of electric light by incandescent lighting lamps then engaged attention. E.A. King in 1845 and W.E. Staite in 1848 had made incandescent electric lamps of an elementary form, and T.A. Edison in 1878 again attacked the problem of producing light by the incandescence of platinum. It had by that time become clear that the most suitable material for an incandescent lamp was carbon contained in a good vacuum, and St G. Lane Fox and Sir J.W. Swan in England, and T.A. Edison in the United States, were engaged in struggling with the difficulties of producing a suitable carbon incandescence electric lamp. Edison constructed in 1879 a successful lamp of this type consisting of a vessel wholly of glass containing a carbon filament made by carbonizing paper or some other carbonizable material, the vessel being exhausted and the current led into the filament through platinum wires. In 1879 and 1880, Edison in the United States, 189
and Swan in conjunction with C.H. Stearn in England, succeeded in completely solving the practical problems. From and after that date incandescent electric lighting became commercially possible, and was brought to public notice chiefly by an electrical exhibition held at the Crystal Palace, near London, in 1882. Edison, moreover, as well as Lane-Fox, had realized the idea of a public electric supply station, and the former proceeded to establish in Pearl Street, New York, in 1881, the first public electric supply station. A similar station in England was opened in the basement of a house in Holborn Viaduct, London, in March 1882. Edison, with copious ingenuity, devised electric meters, electric mains, lamp fittings and generators complete for the purpose. In 1881 C.A. Faure made an important improvement in the lead secondary battery which G. Planté (1834-1889) had invented in 1859, and storage batteries then began to be developed as commercial appliances by Faure, Swan, J.S. Sellon and many others (see Accumulator). In 1882, numerous electric lighting companies were formed for the conduct of public and private lighting, but an electric lighting act passed in that year greatly hindered commercial progress in Great Britain. Nevertheless the delay was utilized in the completion of inventions necessary for the safe and economical distribution of electric current for the purpose of electric lighting.

Telephone.—Going back a few years we find the technical applications of electrical invention had developed themselves in other directions. Alexander Graham Bell in 1876 invented the speaking telephone (*q.v.*), and Edison and Elisha Gray in the United States followed almost immediately with other telephonic inventions for electrically transmitting speech. About the same time D.E. Hughes in England invented the microphone. In 1879 telephone exchanges began to be developed in the United States, Great Britain and other countries.

Electric Power.—Following on the discovery in 1873 of the reversible action of the dynamo and its use as a motor, efforts began to be made to apply this knowledge to transmission of power, and S.D. Field, T.A. Edison, Leo Daft, E.M. Bentley and W.H. Knight, F.J. Sprague, C.J. Van Depoele and others between 1880 and 1884 were the pioneers of electric traction. One of the earliest electric tram cars was exhibited by E.W. and W. Siemens in Paris in 1881. In 1883 Lucien Gaulard, following a line of thought opened by Jablochkov, proposed to employ high pressure alternating currents for electric distributions over wide areas by means of transformers. His ideas were improved by Carl Zipernowsky and O.T. Bláthy in Hungary and by S.Z. de Ferranti in England, and the alternating current transformer (see Transformers) came into existence. Polyphase alternators were first exhibited at the Frankfort electrical exhibition in 1891, developed as a consequence of scientific researches by Galileo Ferraris (1847-1897), Nikola Tesla, M.O. von Dolivo-Dobrowolsky and C.E.L. Brown, and long distance transmission of electrical power by polyphase electrical currents (see Power Transmission: *Electric*) was exhibited in operation at Frankfort in 1891. Meanwhile the early continuous current dynamos devised by Gramme, Siemens and others had been vastly improved in scientific principle and practical construction by the labours of Siemens, J. Hopkinson, R.E.B. Crompton, Elihu Thomson, Rudolf Eickemeyer, Thomas Parker and others, and the theory of the action of the dynamo had been closely studied by J. and E. Hopkinson, G. Kapp, S.P. Thompson, C.P. Steinmetz and J. Swinburne, and great improvements made in the alternating current dynamo by W.M. Mordey, S.Z. de Ferranti and

Messrs Ganz of Budapest. Thus in twenty years from the invention of the Gramme dynamo, electrical engineering had developed from small beginnings into a vast industry. The amendment, in 1888, of the Electric Lighting Act of 1882, before long caused a huge development of public electric lighting in Great Britain. By the end of the 19th century every large city in Europe and in North and South America was provided with a public electric supply for the purposes of electric lighting. The various improvements in electric illuminants, such as the Nernst oxide lamp, the tantalum and osmium incandescent lamps, and improved forms of arc lamp, enclosed, inverted and flame arcs, are described under Lighting: *Electric*.

Between 1890 and 1900, electric traction advanced rapidly in the United States of America but more slowly in England. In 1902 the success of deep tube electric railways in Great Britain was assured, and in 1904 main line railways began to abandon, at least experimentally, the steam locomotive and substitute for it the electric transmission of power. Long distance electrical transmission had been before that time exemplified in the great scheme of utilizing the falls of Niagara. The first projects were discussed in 1891 and 1892 and completed practically some ten years later. In this scheme large turbines were placed at the bottom of hydraulic fall tubes 150 ft. deep, the turbines being coupled by long shafts with 5000 H.P. alternating current dynamos on the surface. By these electric current was generated and transmitted to towns and factories around, being sent overhead as far as Buffalo, a distance of 18 m. At the end of the 19th century electrochemical industries began to be developed which depended on the possession of cheap electric energy. The production of aluminium in Switzerland and Scotland, carborundum and calcium carbide in the United States, and soda by the Castner-Kellner process, began to be conducted on an immense scale. The early work of Sir W. Siemens on the electric furnace was continued and greatly extended by Henri Moissan and others on its scientific side, and electrochemistry took its place as one of the most promising departments of technical research and invention. It was stimulated and assisted by improvements in the construction of large dynamos and increased knowledge concerning the control of powerful electric currents.

In the early part of the 20th century the distribution in bulk of electric energy for power purposes in Great Britain began to assume important proportions. It was seen to be uneconomical for each city and town to manufacture its own supply since, owing to the intermittent nature of the demand for current for lighting, the price had to be kept up to 4d. and 6d. per unit. It was found that by the manufacture in bulk, even by steam engines, at primary centres the cost could be considerably reduced, and in numerous districts in England large power stations began to be erected between 1903 and 1905 for the supply of current for power purposes. This involved almost a revolution in the nature of the tools used, and in the methods of working, and may ultimately even greatly affect the factory system and the concentration of population in large towns which was brought about in the early part of the 19th century by the invention of the steam engine.

Development of Electric Theory.

Turning now to the theory of electricity, we may note the equally remarkable progress made in 300 years in scientific insight into the nature of the agency which has so recast the face of human society. There is no need to dwell upon the early crude theories of the action of amber and lodestone. In a true scientific sense no hypothesis was possible, because few facts had been accumulated. The discoveries of Stephen Gray and C.F. de C. du Fay on the conductivity of some bodies for the electric agency and the dual character of electrification gave rise to the first notions of electricity as an imponderable fluid, or non-gravitative subtile matter, of a more refined and penetrating kind than ordinary liquids and gases. Its duplex character, and the fact that the electricity produced by rubbing glass and vitreous substances was different from that produced by rubbing sealing-wax and resinous substances, seemed to necessitate the assumption of two kinds of electric fluid; hence there arose the conception of *positive* and *negative* electricity, and the two-fluid theory came into existence.

Single-fluid Theory.—The study of the phenomena of the Leyden jar and of the fact that the inside and outside coatings possessed opposite electricities, so that in charging the jar as much positive electricity is added to one side as negative to the other, led Franklin about 1750 to suggest a modification called the single fluid theory, in which the two states of electrification were regarded as not the results of two entirely 190 different fluids but of the addition or subtraction of one electric fluid from matter, so that positive electrification was to be looked upon as the result of increase or addition of something to ordinary matter and negative as a subtraction. The positive and negative electrifications of the two coatings of the Leyden jar were therefore to be regarded as the result of a transformation of something called electricity from one coating to the other, by which process a certain measurable quantity became so much less on one side by the same amount by which it became more on the other. A modification of this single fluid theory was put forward by F.U.T. Aepinus which was explained and illustrated in his *Tentamen theoriae electricitatis et magnetismi*, published in St Petersburg in 1759. This theory was founded on the following principles:—(1) the particles of the electric fluid repel each other with a force decreasing as the distance increases; (2) the particles of the electric fluid attract the atoms of all bodies and are attracted by them with a force obeying the same law; (3) the electric fluid exists in the pores of all bodies, and while it moves without any obstruction in conductors such as metals, water, &c., it moves with extreme difficulty in so-called non-conductors such as glass, resin, &c.; (4) electrical phenomena are produced either by the transference of the electric fluid of a body containing more to one containing less, or from its attraction and repulsion when no transference takes place. Electric attractions and repulsions were, however, regarded as differential actions in which the mutual repulsion of the particles of electricity operated, so to speak, in antagonism to the mutual attraction of particles of matter for one another and of particles of electricity for matter. Independently of Aepinus, Henry Cavendish put forward a single-fluid theory of electricity (*Phil. Trans.*, 1771, 61, p. 584), in which he considered it in more precise detail.

Two-fluid Theory.—In the elucidation of electrical phenomena, however, towards the end of the 18th century, a modification of the two-fluid theory seems to have been generally preferred. The notion then formed of the nature of electrification was something as follows:—All bodies were assumed to contain a certain quantity of a so-called neutral fluid made up of equal quantities of positive and negative electricity, which when in this state of combination neutralized one another's properties. The neutral fluid could, however, be divided up or separated into its two constituents, and these could be accumulated on separate conductors or non-conductors. This view followed from the discovery of the facts of electric induction of J. Canton (1753, 1754). When, for instance, a positively electrified body was found to induce upon another insulated conductor a charge of negative electricity on the side nearest to it, and a charge of positive electricity on the side farthest from it, this was explained by saying that the particles of each of the two electric fluids repelled one another but attracted those of the positive fluid. Hence the operation of the positive charge upon the neutral fluid was to draw towards the positive the negative constituent of the neutral charge and repel to the distant parts of the conductor the positive constituent.

C.A. Coulomb experimentally proved that the law of attraction and repulsion of simple electrified bodies was that the force between them varied inversely as the square of the distance and thus gave mathematical definiteness to the two-fluid hypothesis. It was then assumed that each of the two constituents of the neutral fluid had an atomic structure and that the so-called particles of one of the electric fluids, say positive, repelled similar particles with a force varying inversely as a square of the distance and attracted those of the opposite fluid according to the same law. This fact and hypothesis brought electrical phenomena within the domain of mathematical analysis and, as already mentioned, Laplace, Biot, Poisson, G.A.A. Plana (1781-1846), and later Robert Murphy (1806-1843), made them the subject of their investigations on the mode in which electricity distributes itself on conductors when in equilibrium.

Faraday's Views.—The two-fluid theory may be said to have held the field until the time when Faraday began his researches on electricity. After he had educated himself by the study of the phenomena of lines of magnetic force in his discoveries on electromagnetic induction, he applied the same conception to electrostatic phenomena, and thus created the notion of lines of electrostatic force and of the important function of the dielectric or non-conductor in sustaining them. Faraday's notion as to the nature of electrification, therefore, about the middle of the 19th century came to be something as follows:—He considered that the so-called charge of electricity on a conductor was in reality nothing on the conductor or in the conductor itself, but consisted in a state of strain or polarization, or a physical change of some kind in the particles of the dielectric surrounding the conductor, and that it was this physical state in the dielectric which constituted electrification. Since Faraday was well aware that even a good vacuum can act as a dielectric, he recognized that the state he called dielectric polarization could not be wholly dependent upon the presence of gravitative matter, but that there must be an electromagnetic medium of a supermaterial nature. In the 13th series of his *Experimental Researches on Electricity* he discussed the relation of a vacuum to electricity. Furthermore his electrochemical investigations, and particularly his discovery of the important law of electrolysis, that the movement of a certain quantity of electricity through an electrolyte is always accompanied by the transfer of a certain definite quantity of matter from one electrode to another and the liberation at these electrodes of an equivalent weight of the ions, gave foundation for the idea of a definite atomic charge of electricity. In fact, long previously to Faraday's electrochemical researches, Sir H. Davy and J.J. Berzelius early in the 19th century had advanced the hypothesis that chemical combination was due to electric attractions between the electric charges carried by chemical atoms. The notion, however, that electricity is atomic in structure was definitely put forward by Hermann von Helmholtz in a well-known Faraday lecture. Helmholtz says: "If we accept the hypothesis that elementary substances are composed of atoms, we cannot well avoid concluding that electricity also is divided into elementary portions which behave like atoms of electricity."[16] Clerk Maxwell had already used in 1873 the phrase, "a molecule of electricity."[17] Towards the end of the third quarter of the 19th century it therefore became clear that electricity, whatever be its nature, was associated with atoms of matter in the form of exact multiples of an indivisible minimum electric charge which may be considered to be "Nature's unit of electricity." This ultimate unit of electric quantity Professor Johnstone Stoney called an *electron*.[18] The formulation of electrical theory as far as regards operations in space free from matter was immensely assisted by Maxwell's mathematical theory. Oliver Heaviside after 1880 rendered much assistance by reducing Maxwell's mathematical analysis to more compact form and by introducing greater precision into terminology (see his *Electrical Papers*, 1892). This is perhaps the place to refer also to the great services of Lord Rayleigh to electrical science. Succeeding Maxwell as Cavendish professor of physics at Cambridge in 1880, he soon devoted himself especially to the exact redetermination of the practical electrical units in absolute measure. He followed up the early work of the British Association Committee on electrical units by a fresh determination of the ohm in absolute measure, and in conjunction with other work on the electrochemical equivalent of silver and the absolute electromotive force of the Clark cell may be said to have placed exact electrical measurement on a new basis. He also made great additions to the theory of alternating electric currents, and provided fresh appliances for other electrical measurements (see his *Collected Scientific Papers*, Cambridge, 1900).

Electro-optics.—For a long time Faraday's observation on the rotation of the plane of polarized light by heavy glass in a magnetic field remained an isolated fact in electro-optics. Then M.E. Verdet (1824-1860) 191 made a study of the subject and discovered that a solution of ferric perchloride in methyl alcohol rotated the plane of polarization in an opposite direction to heavy glass (*Ann. Chim. Phys.*, 1854, 41, p. 370; 1855, 43, p. 37; *Com. Rend.*, 1854, 39, p. 548). Later A.A.E.E. Kundt prepared metallic films of iron, nickel and cobalt, and obtained powerful negative optical rotation with them (*Wied. Ann.*, 1884, 23, p. 228; 1886, 27, p. 191). John Kerr (1824-1907) discovered that a similar effect was produced when plane polarized light was reflected from the pole of a powerful magnet (*Phil. Mag.*, 1877, [5], 3, p. 321, and 1878, 5, p. 161). Lord Kelvin showed that Faraday's discovery demonstrated that some form of rotation was taking place along lines of magnetic force when passing through a medium.[19] Many observers have given attention to the exact determination of Verdet's constant of rotation for standard substances, *e.g.* Lord Rayleigh for carbon bisulphide,[20] and Sir W.H. Perkin for an immense range of inorganic and organic bodies.[21] Kerr also discovered that when certain homogeneous dielectrics were submitted to electric strain, they became birefringent (*Phil. Mag.*, 1875, 50, pp. 337 and 446). The theory of electro-optics received great attention from Kelvin, Maxwell, Rayleigh, G.F. Fitzgerald, A. Righi and P.K.L. Drude, and experimental contributions from innumerable workers, such as F.T. Trouton, O.J. Lodge and J.L. Howard, and many others.

Electric Waves.—In the decade 1880-1890, the most important advance in electrical physics was, however, that which originated with the astonishing researches of Heinrich Rudolf Hertz (1857-1894). This illustrious investigator was stimulated, by a certain problem brought to his notice by H. von Helmholtz, to undertake investigations which had for their object a demonstration of the truth of Maxwell's principle that a variation in electric displacement was in fact an electric current and had magnetic effects. It is impossible to describe here the details of these elaborate experiments; the reader must be referred to Hertz's own papers, or the English translation of them by Prof. D.E. Jones. Hertz's great discovery was an experimental realization of a suggestion made by G.F. Fitzgerald (1851-1901) in 1883 as to a method of producing electric waves in space. He invented for this purpose a radiator consisting of two metal rods placed in one line, their inner ends being provided with poles nearly touching and their outer ends with metal plates. Such an arrangement constitutes in effect a condenser, and when the two plates respectively are connected to the secondary terminals of an induction coil in operation, the plates are rapidly and alternately charged, and discharged across the spark gap with electrical oscillations (see Electrokinetics). Hertz then devised a wave detecting apparatus called a resonator. This in its simplest form consisted of a ring of wire nearly closed terminating in spark balls very close together, adjustable as to distance by a micrometer screw. He found that when the resonator was placed in certain positions with regard to the oscillator, small sparks were seen between the micrometer balls, and when the oscillator was placed at one end of a room having a sheet of zinc fixed against the wall at the other end, symmetrical positions could be found in the room at which, when the resonator was there placed, either no sparks or else very bright sparks occurred at the poles. These effects, as Hertz showed, indicated the establishment of stationary electric waves in space and the propagation of electric and magnetic force through space with a finite velocity. The other additional phenomena he observed finally contributed an all but conclusive proof of the truth of Maxwell's views. By profoundly ingenious methods Hertz showed that these invisible electric waves could be reflected and refracted like waves of light by mirrors and prisms, and that familiar experiments in optics could be repeated with electric waves which could not affect the eye. Hence there arose a new science of electro-optics, and in all parts of Europe and the United States innumerable investigators took possession of the novel field of research with the greatest delight. O.J. Lodge,[22] A. Righi,[23] J.H. Poincaré,[24] V.F.K. Bjerknes, P.K.L. Drude, J.J. Thomson,[25] John Trowbridge, Max Abraham, and many others, contributed to its elucidation.

In 1892, E. Branly of Paris devised an appliance for detecting these waves which subsequently proved to be of immense importance. He discovered that they had the power of affecting the electric conductivity of materials when in a state of powder, the majority of metallic filings increasing in conductivity. Lodge devised a similar arrangement called a coherer, and E. Rutherford invented a magnetic detector depending on the power of electric oscillations to demagnetize iron or steel. The sum total of all these contributions to electrical knowledge had the effect of establishing Maxwell's principles on a firm basis, but they also led to technical inventions of the very greatest utility. In 1896 G. Marconi applied a modified and improved form of Branly's wave detector in conjunction with a novel form of radiator for the telegraphic transmission of intelligence through space without wires, and he and others developed this new form of telegraphy with the greatest rapidity and success into a startling and most useful means of communicating through space electrically without connecting wires.

Electrolysis.—The study of the transfer of electricity through liquids had meanwhile received much attention. The general facts and laws of electrolysis (*q.v.*) were determined experimentally by Davy and Faraday and confirmed by the researches of J.F. Daniell, R.W. Bunsen and Helmholtz. The modern theory of electrolysis grew up under the hands of R.J.E. Clausius, A.W. Williamson and F.W.G. Kohlrausch, and received a great impetus from the work of Svante Arrhenius, J.H. Van't Hoff, W. Ostwald, H.W. Nernst and many others. The theory of the ionization of salts in solution has raised much discussion amongst chemists, but the general fact is certain that electricity only moves through liquids in association with matter, and simultaneously involves chemical dissociation of molecular groups.

Discharge through Gases.—Many eminent physicists had an instinctive feeling that the study of the passage of electricity through gases would shed much light on the intrinsic nature of electricity. Faraday devoted to a careful examination of the phenomena the XIIIth series of his *Experimental Researches*, and among the older workers in this field must be particularly mentioned J. Plücker, J.W. Hittorf, A.A. de la Rive, J.P. Gassiot, C.F. Varley, and W. Spottiswoode and J. Fletcher Moulton. It has long been known that air and other gases at the pressure of the atmosphere were very perfect insulators, but that when they were rarefied and contained in glass tubes with platinum electrodes sealed through the glass, electricity could be passed through them under sufficient electromotive force and produced a luminous appearance known as the electric glow discharge. The so-called vacuum tubes constructed by H. Geissler (1815-1879) containing air, carbonic acid, hydrogen, &c., under a pressure of one or two millimetres, exhibit beautiful appearances when traversed by the high tension current produced by the secondary circuit of an induction coil. Faraday discovered the existence of a dark space round the negative electrode which is usually known as the "Faraday dark space." De la Rive added much to our knowledge of the subject, and J. Plücker and his disciple J.W. Hittorf examined the phenomena exhibited in so-called high vacua, that is, in exceedingly rarefied gases. C.F. Varley discovered the interesting fact that no current could be sent through the rarefied gas unless a certain minimum potential difference of the electrodes was excited. Sir William Crookes took up in 1872 the study of electric discharge through high vacua, having been led to it by his researches on the radiometer. 192
The particular details of the phenomena observed will be found described in the article Conduction, Electric (§ III.). The main fact discovered by researches of Plücker, Hittorf and Crookes was that in a vacuum tube containing extremely rarefied air or other gas, a luminous discharge takes place from the negative electrode which proceeds in lines normal to the surface of the negative electrode and renders phosphorescent both the glass envelope and other objects placed in the vacuum tube when it falls upon them. Hittorf made in 1869 the discovery that solid objects could cast shadows or intercept this cathode discharge. The cathode discharge henceforth engaged the attention of many physicists. Varley had advanced tentatively the hypothesis that it consisted in an actual projection of electrified matter from the cathode, and Crookes was led by his researches in 1870, 1871 and 1872 to embrace and confirm this hypothesis in a modified form and announce the existence of a fourth state of matter, which he called radiant matter, demonstrating by many beautiful and convincing experiments that there was an actual projection of material substance of some kind

possessing inertia from the surface of the cathode. German physicists such as E. Goldstein were inclined to take another view. Sir J.J. Thomson, the successor of Maxwell and Lord Rayleigh in the Cavendish chair of physics in the university of Cambridge, began about the year 1899 a remarkable series of investigations on the cathode discharge, which finally enabled him to make a measurement of the ratio of the electric charge to the mass of the particles of matter projected from the cathode, and to show that this electric charge was identical with the atomic electric charge carried by a hydrogen ion in the act of electrolysis, but that the mass of the cathode particles, or "corpuscles" as he called them, was far less, viz. about $\frac{1}{2000}$th part of the mass of a hydrogen atom.[26] The subject was pursued by Thomson and the Cambridge physicists with great mathematical and experimental ability, and finally the conclusion was reached that in a high vacuum tube the electric charge is carried by particles which have a mass only a fraction, as above mentioned, of that of the hydrogen atom, but which carry a charge equal to the unit electric charge of the hydrogen ion as found by electrochemical researches.[27] P.E.A. Lenard made in 1894 (*Wied. Ann. Phys.*, 51, p. 225) the discovery that these cathode particles or corpuscles could pass through a window of thin sheet aluminium placed in the wall of the vacuum tube and give rise to a class of radiation called the Lenard rays. W.C. Röntgen of Munich made in 1896 his remarkable discovery of the so-called X or Röntgen rays, a class of radiation produced by the impact of the cathode particles against an impervious metallic screen or anticathode placed in the vacuum tube. The study of Röntgen rays was ardently pursued by the principal physicists in Europe during the years 1897 and 1898 and subsequently. The principal property of these Röntgen rays which attracted public attention was their power of passing through many solid bodies and affecting a photographic plate. Hence some substances were opaque to them and others transparent. The astonishing feat of photographing the bones of the living animal within the tissues soon rendered the Röntgen rays indispensable in surgery and directed an army of investigators to their study.

Radioactivity.—One outcome of all this was the discovery by H. Becquerel in 1896 that minerals containing uranium, and particularly the mineral known as pitchblende, had the power of affecting sensitive photographic plates enclosed in a black paper envelope when the mineral was placed on the outside, as well as of discharging a charged electroscope (*Com. Rend.*, 1896, 122, p. 420). This research opened a way of approach to the phenomena of radioactivity, and the history of the steps by which P. Curie and Madame Curie were finally led to the discovery of radium is one of the most fascinating chapters in the history of science. The study of radium and radioactivity (see Radioactivity) led before long to the further remarkable knowledge that these so-called radioactive materials project into surrounding space particles or corpuscles, some of which are identical with those projected from the cathode in a high vacuum tube, together with others of a different nature. The study of radioactivity was pursued with great ability not only by the Curies and A. Debierne, who associated himself with them, in France, but by E. Rutherford and F. Soddy in Canada, and by J.J. Thomson, Sir William Crookes, Sir William Ramsay and others in England.

Electronic Theory.—The final outcome of these investigations was the hypothesis that Thomson's corpuscles or particles composing the cathode discharge in a high vacuum tube must be looked upon as the ultimate constituent of what we call negative electricity; in other words, they are atoms of negative electricity, possessing, however, inertia, and these negative electrons are components at any rate of the chemical atom. Each electron is a point-charge of negative electricity equal to 3.9×10^{-10} of an electrostatic unit or to 1.3×10^{-20} of an electromagnetic unit, and the ratio of its charge to its mass is nearly 2×10^{7} using E.M. units. For the hydrogen atom the ratio of charge to mass as deduced from electrolysis is about 10^{4}. Hence the mass of an electron is $\frac{1}{2000}$th of that of a hydrogen atom. No one has yet been able to isolate positive electrons, or to give a complete demonstration that the whole inertia of matter is only electric inertia due to what may be called the inductance of the electrons. Prof. Sir J. Larmor developed in a series of very able papers (*Phil. Trans.*, 1894, 185; 1895, 186; 1897, 190), and subsequently in his book *Aether and Matter* (1900), a remarkable hypothesis of the structure of the electron or corpuscle, which he regards as simply a strain centre in the aether or electromagnetic medium, a chemical atom being a collection of positive and negative electrons or strain centres in stable orbital motion round their common centre of mass (see Aether). J.J. Thomson also developed this hypothesis in a profoundly interesting manner, and we may therefore summarize very briefly the views held on the nature of electricity and matter at the beginning of the 20th century by saying that the term electricity had come to be regarded, in part at least, as a collective name for electrons, which in turn must be considered as constituents of the chemical atom, furthermore as centres of certain lines of self-locked and permanent strain existing in the universal aether or electromagnetic medium. Atoms of matter are composed of congeries of electrons and the inertia of matter is probably therefore only the inertia of the electromagnetic medium.[28] Electric waves are produced wherever electrons are accelerated or retarded, that is, whenever the velocity of an electron is changed or accelerated positively or negatively. In every solid body there is a continual atomic dissociation, the result of which is that mixed up with the atoms of chemical matter composing them we have a greater or less percentage of free electrons. The operation called an electric current consists in a diffusion or movement of these electrons through matter, and this is controlled by laws of diffusion which are similar to those of the diffusion of liquids or gases. Electromotive force is due to a difference in the density of the electronic population in different or identical conducting bodies, and whilst the electrons can move freely through so-called conductors their motion is much more hindered or restricted in non-conductors. Electric charge consists, therefore, in an excess or deficit of negative electrons in a body. In the hands of H.A. Lorentz, P.K.L. Drude, J. J, Thomson, J. Larmor and many others, the electronic hypothesis of matter and of electricity has been developed in great detail and may be said to represent the outcome of modern researches upon electrical phenomena.

The reader may be referred for an admirable summary of the theories of electricity prior to the advent of 193
the electronic hypothesis to J.J. Thomson's "Report on Electrical Theories" (*Brit. Assoc. Report*, 1885), in which he divides electrical theories enunciated during the 19th century into four classes, and summarizes the opinions and theories of A.M. Ampère, H.G. Grassman, C.F. Gauss, W.E. Weber, G.F.B. Riemann, R.J.E. Clausius, F.E. Neumann and H. von Helmholtz.

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1890); H.M. Noad, *A Manual of Electricity* (2 vols., London, 1855, 1857); J.J. Thomson, *Recent Researches in Electricity and Magnetism* (Oxford, 1893); id., *Conduction of Electricity through Gases* (Cambridge, 1903); id., *Electricity and Matter* (London, 1904); O. Heaviside, *Electromagnetic Theory* (London, 1893); O.J. Lodge, *Modern Views of Electricity* (London, 1889); E. Mascart and J. Joubert, *A Treatise on Electricity and Magnetism*, English trans. by E. Atkinson (2 vols., London, 1883); Park Benjamin, *The Intellectual Rise in Electricity* (London, 1895); G.C. Foster and A.W. Porter, *Electricity and Magnetism* (London, 1903); A. Gray, *A Treatise on Magnetism and Electricity* (London, 1898); H.W. Watson and S.H. Burbury, *The Mathematical Theory of Electricity and Magnetism* (2 vols., 1885); Lord Kelvin (Sir William Thomson), *Mathematical and Physical Papers* (3 vols., Cambridge, 1882); Lord Rayleigh, *Scientific Papers* (4 vols., Cambridge, 1903); A. Winkelmann, *Handbuch der Physik*, vols. iii. and iv. (Breslau, 1903 and 1905; a mine of wealth for references to original papers on electricity and magnetism from the earliest date up to modern times). For particular information on the modern Electronic theory the reader may consult W. Kaufmann, "The Developments of the Electron Idea." *Physikalische Zeitschrift* (1st of Oct. 1901), or *The Electrician* (1901), 48, p. 95; H.A. Lorentz, *The Theory of Electrons* (1909); E.E. Fournier d'Albe, *The Electron Theory* (London, 1906); H. Abraham and P. Langevin, *Ions, Electrons, Corpuscles* (Paris, 1905); J.A. Fleming, "The Electronic Theory of Electricity," *Popular Science Monthly* (May 1902); Sir Oliver J. Lodge, *Electrons, or the Nature and Properties of Negative Electricity* (London, 1907).

(J. A. F.)

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1 Gilbert's work, *On the Magnet, Magnetic Bodies and the Great Magnet, the Earth*, has been translated from the rare folio Latin edition of 1600, but otherwise reproduced in its original form by the chief members of the Gilbert Club of England, with a series of valuable notes by Prof. S.P. Thompson (London, 1900). See also *The Electrician*, February 21, 1902.

2 See *The Intellectual Rise in Electricity*, ch. x., by Park Benjamin (London, 1895).

3 See Sir Oliver Lodge, "Lightning, Lightning Conductors and Lightning Protectors," *Journ. Inst. Elec. Eng.* (1889), 18, p. 386, and the discussion on the subject in the same volume; also the book by the same author on *Lightning Conductors and Lightning Guards* (London, 1892).

4 *The Electrical Researches of the Hon. Henry Cavendish 1771-1781*, edited from the original manuscripts by J. Clerk Maxwell, F.R.S. (Cambridge, 1879).

5 In 1878 Clerk Maxwell repeated Cavendish's experiments with improved apparatus and the employment of a Kelvin quadrant electrometer as a means of detecting the absence of charge on the inner conductor after it had been connected to the outer case, and was thus able to show that if the law of electric attraction varies inversely as the nth power of the distance, then the exponent n must have a value of $2 \pm \frac{1}{21600}$. See Cavendish's *Electrical Researches*, p. 419.

6 Modern researches have shown that the loss of charge is in fact dependent upon the ionization of the air, and that, provided the atmospheric moisture is prevented from condensing on the insulating supports, water vapour in the air does not *per se* bestow on it conductance for electricity.

7 Faraday discussed the chemical theory of the pile and arguments in support of it in the 8th and 16th series of his *Experimental Researches on Electricity*. De la Rive reviews the subject in his large *Treatise on Electricity and Magnetism*, vol. ii. ch. iii. The writer made a contribution to the discussion in 1874 in a paper on "The Contact Theory of the Galvanic Cell," *Phil. Mag.*, 1874, 47, p. 401. Sir Oliver Lodge reviewed the whole position in a paper in 1885, "On the Seat of the Electromotive Force in a Voltaic Cell," *Journ. Inst. Elec. Eng.*, 1885, 14, p. 186.

8 "Mémoire sur la théorie mathématique des phénomènes électrodynamiques," *Mémoires de l'institut*, 1820, 6; see also *Ann. de Chim.*, 1820, 15.

9 See M. Faraday, "On some new Electro-Magnetical Motions and on the Theory of Magnetism," *Quarterly Journal of Science*, 1822, 12, p. 74; or *Experimental Researches on Electricity*, vol. ii. p. 127.

10 Amongst the most important of Faraday's quantitative researches must be included the ingenious and convincing proofs he provided that the production of any quantity of electricity of one sign is always accompanied by the production of an equal quantity of electricity of the opposite sign. See *Experimental Researches on Electricity*, vol. i. § 1177.

11 In this connexion the work of George Green (1793-1841) must not be forgotten. Green's *Essay on the Application of Mathematical Analysis to the Theories of Electricity and Magnetism*, published in 1828, contains the first exposition of the theory of potential. An important theorem contained in it is known as Green's theorem, and is of great value.

12 See also his *Submarine Telegraphs* (London, 1898).

13 The quantitative study of electrical phenomena has been enormously assisted by the establishment of the absolute system of electrical measurement due originally to Gauss and Weber. The British Association for the advancement of science appointed in 1861 a committee on electrical units, which made its first report in 1862 and has existed ever since. In this work Lord Kelvin took a leading part. The popularization of the system was greatly assisted by the publication by Prof. J.D. Everett of *The C.G.S. System of Units* (London, 1891).

14 The first paper in which Maxwell began to translate Faraday's conceptions into mathematical language was "On Faraday's Lines of Force," read to the Cambridge Philosophical Society on the 10th of December 1855 and the 11th of February 1856. See Maxwell's *Collected Scientific Papers*, i. 155.

15 *A Treatise on Electricity and Magnetism* (2 vols.), by James Clerk Maxwell, sometime professor of experimental physics in the university of Cambridge. A second edition was edited by Sir W.D. Niven in 1881 and a third by Prof. Sir J.J. Thomson in 1891.

16 H. von Helmholtz, "On the Modern Development of Faraday's Conception of Electricity," *Journ. Chem. Soc.*, 1881, 39, p. 277.

17 See Maxwell's *Electricity and Magnetism*, vol. i. p. 350 (2nd ed., 1881).

18 "On the Physical Units of Nature," *Phil. Mag.*, 1881, [5], 11, p. 381. Also *Trans. Roy. Soc.* (Dublin, 1891), 4, p. 583.

19 See Sir W. Thomson, *Proc. Roy. Soc. Lond.*, 1856, 8, p. 152; or Maxwell, *Elect. and Mag.*, vol. ii. p. 831.

20 See Lord Rayleigh, *Proc. Roy. Soc. Lond.*, 1884, 37, p. 146; Gordon, *Phil. Trans.*, 1877, 167, p. 1; H. Becquerel,

Ann. Chim. Phys., 1882, [3], 27, p. 312.

21 Perkin's Papers are to be found in the *Journ. Chem. Soc. Lond.*, 1884, p. 421; 1886, p. 177; 1888, p. 561; 1889, p. 680; 1891, p. 981; 1892, p. 800; 1893, p. 75.

22 *The Work of Hertz* (London, 1894).

23 *L'Ottica delle oscillazioni elettriche* (Bologna, 1897).

24 *Les Oscillations électriques* (Paris, 1894).

25 *Recent Researches in Electricity and Magnetism* (Oxford, 1892).

26 See J.J. Thomson, *Proc. Roy. Inst. Lond.*, 1897, 15, p. 419; also *Phil. Mag.*, 1899, [5], 48, p. 547.

27 Later results show that the mass of a hydrogen atom is not far from 1.3×10^{-24} gramme and that the unit atomic charge or natural unit of electricity is 1.3×10^{-20} of an electromagnetic C.G.S. unit. The mass of the electron or corpuscle is 7.0×10^{-28} gramme and its diameter is 3×10^{-13} centimetre. The diameter of a chemical atom is of the order of 10^{-7} centimetre.

See H.A. Lorentz, "The Electron Theory," *Elektrotechnische Zeitschrift*, 1905, 26, p. 584; or *Science Abstracts*, 1905, 8, A, p. 603.

28 See J.J. Thomson, *Electricity and Matter* (London, 1904).

ELECTRICITY SUPPLY. I. *General Principles.*—The improvements made in the dynamo and electric motor between 1870 and 1880 and also in the details of the arc and incandescent electric lamp towards the close of that decade, induced engineers to turn their attention to the question of the private and public supply of electric current for the purpose of lighting and power. T.A. Edison[1] and St G. Lane Fox[2] were among the first to see the possibilities and advantages of public electric supply, and to devise plans for its practical establishment. If a supply of electric current has to be furnished to a building the option exists in many cases of drawing from a public supply or of generating it by a private plant.

Private Plants.—In spite of a great amount of ingenuity devoted to the development of the primary battery and the thermopile, no means of generation of large currents can compete in economy with the dynamo. Hence a private electric generating plant involves the erection of a dynamo which may be driven either by a steam, gas or oil engine, or by power obtained by means of a turbine from a low or high fall of water. It may be either directly coupled to the motor, or driven by a belt; and it may be either a continuous-current machine or an alternator, and if the latter, either single-phase or polyphase. The convenience of being able to employ storage batteries in connexion with a private-supply system is so great that unless power has to be transmitted long distances, the invariable rule is to employ a continuous-current dynamo. Where space is valuable this is always coupled direct to the motor; and if a steam-engine is employed, an enclosed engine is most cleanly and compact. Where coal or heating gas is available, a gas-engine is exceedingly convenient, since it requires little attention. Where coal gas is not available, a Dowson gas-producer can be employed. The oil-engine has been so improved that it is extensively used in combination with a direct-coupled or belt-driven dynamo and thus forms a favourite and easily-managed plant for private electric lighting. Lead storage cells, however, as at present made, when charged by a steam-driven dynamo deteriorate less rapidly than when an oil-engine is employed, the reason being that the charging current is more irregular in the latter case, since the single cylinder oil-engine only makes an impulse every other revolution. In connexion with the generator, it is almost the invariable custom to put down a secondary battery of storage cells, to enable the supply to be given after the engine has stopped. This is necessary, not only as a security for the continuity of supply, but because otherwise the costs of labour in running the engine night and day become excessive. The storage battery gives its supply automatically, but the dynamo and engine require incessant skilled attendance. If the building to be lighted is at some distance from the engine-house the battery should be placed in the basement of the building, and underground or overhead conductors, to convey the charging current, brought to it from the dynamo.

It is usual, in the case of electric lighting installations, to reckon all lamps in their equivalent number of 8 candle power (c.p.) incandescent lamps. In lighting a private house or building, the first thing to be done is to settle the total number of incandescent lamps and their size, whether 32 c.p., 16 c.p. or 8 c.p. Lamps of 5 c.p. can be used with advantage in small bedrooms and passages. Each candle-power in the case of a carbon filament lamp can be taken as equivalent to 3.5 watts, or the 8 c.p. lamp as equal to 30 watts, the 16 c.p. lamp to 60 watts, and so on. In the case of metallic filament lamps about 1.0 or 1.25 watts. Hence if the equivalent of 100 carbon filament 8 c.p. lamps is required in a building the maximum electric power-supply available must be 3000 watts or 3 kilowatts. The next matter to consider is the pressure of supply. If the battery can be in a position near the building to be lighted, it is best to use 100-volt incandescent lamps and enclosed arc lamps, which can be worked singly off the 100-volt circuit. If, however, the lamps are scattered over a wide area, or in separate buildings somewhat far apart, as in a college or hospital, it may be better to select 200 volts as the supply pressure. Arc lamps can then be worked three in series with added resistance. The third step is to select the size of the dynamo unit and the amount of spare plant. It is desirable that there should be at least three dynamos, two of which are capable of taking the whole of the full load, the third being reserved to replace either of the others when required. The total power to be absorbed by the lamps and motors (if any) being given, together with an allowance for extensions, the size of the dynamos can be settled, and the power of the engines required to drive them determined. A good rule to follow is that the indicated horse-power (I.H.P.) of the engine should be double the dynamo full-load output in kilowatts; that is to say, for a 10-kilowatt dynamo an engine should be capable of giving 20 indicated (not nominal) H.P. From the I.H.P. of the engine, if a steam engine, the size of the boiler required for steam production becomes known. For small plants it is safe to reckon that, including water waste, boiler capacity should be provided equal to evaporating 40 ℔ of water per hour for every I.H.P. of the engine. The locomotive boiler is a convenient form; but where large amounts of steam are required, some modification of the Lancashire boiler or the water-tube boiler is generally adopted. In settling the electromotive force of the dynamo to be employed, attention must be paid to the question of charging secondary cells, if these are used. If a secondary battery is employed in connexion with 100-volt lamps, it is usual to put in 53 or 54 cells. The electromotive force of these cells varies between 2.2 and 1.8 volts as they discharge; hence the above number of cells is sufficient for maintaining the necessary electromotive force. For charging, however, it is necessary to provide 2.5 volts per cell, and the dynamo must therefore have an electromotive force of 135 volts, *plus* any voltage required to overcome the fall of potential in the cable connecting the dynamo with the secondary battery. Supposing this to be 10 volts, it is safe to install dynamos having an electromotive force of 150 volts, since by means of resistance in the field circuits this electromotive force can be lowered to 110 or 115 if it is required at any time to dispense with the battery. The size of the secondary cell will be determined by the nature of the supply to be given after the dynamos have been stopped. It is usual to provide sufficient 194 storage capacity to run all the lamps for three or four hours without assistance from the dynamo.

As an example taken from actual practice, the following figures give the capacity of the plant put down to supply 500 8 c.p. lamps in a hospital. The dynamos were 15-unit machines, having a full-load capacity of 100 amperes at 150 volts, each coupled direct to an engine of 25 H.P.; and a double plant of this description was supplied from two steel locomotive boilers, each capable of evaporating 800 ℔ of water per hour. One dynamo during the day was used for charging the storage battery of 54 cells; and at night the discharge from the cells, together with the current from one of the dynamos, supplied the lamps until the heaviest part of the load had been taken; after that the current was drawn from the batteries alone. In working such a plant it is necessary to have the means of varying the electromotive force of the dynamo as the charging of the cells proceeds. When they are nearly exhausted, their electromotive force is less than 2 volts; but as the charging proceeds, a counter-electromotive force is gradually built up, and the engineer-in-charge has to raise the voltage of the dynamo in order to maintain a constant charging current. This is effected by having the dynamos designed to give normally the highest E.M.F. required, and then inserting resistance in their field circuits to reduce it as may be necessary. The space and attendance required for an oil-engine plant are much less than for a steam-engine.

Public Supply.—The methods at present in successful operation for public electric supply fall into two broad divisions:—(1) continuous-current systems and (2) alternating-current systems. Continuous-current systems are either low- or high-pressure. In the former the current is generated by dynamos at some pressure less than 500 volts, generally about 460 volts, and is supplied to users at half this pressure by means of a three-wire system (see below) of distribution, with or without the addition of storage batteries.

Low-pressure continuous supply.

The general arrangements of a low-pressure continuous-current town supply station are as follows:—If steam is the motive power selected, it is generated under all the best conditions of economy by a battery of boilers, and supplied to engines which are now almost invariably coupled direct, each to its own dynamo, on one common bedplate; a multipolar dynamo is most usually employed, coupled direct to an enclosed engine. Parsons or Curtis steam turbines (see Steam-Engine) are frequently selected, since experience has shown that the costs of oil and attendance are far less for this type than for the reciprocating engine, whilst the floor space and, therefore, the building cost are greatly reduced. In choosing the size of unit to be adopted, the engineer has need of considerable experience and discretion, and also a full knowledge of the nature of the public demand for electric current. The rule is to choose as large units as possible, consistent with security, because they are proportionately more economical than small ones. The over-all efficiency of a steam dynamo—that is, the ratio between the electrical power output, reckoned say in kilowatts, and the I.H.P. of the engine, reckoned in the same units—is a number which falls rapidly as the load decreases, but at full load may reach some such value as 80 or 85%. It is common to specify the efficiency, as above defined, which must be attained by the plant at full-load, and also the efficiencies at quarter- and half-load which must be reached or exceeded. Hence in the selection of the size of the units the engineer is guided by the consideration that whatever units are in use shall be as nearly as possible fully loaded. If the demand on the station is chiefly for electric lighting, it varies during the hours of the day and night with tolerable regularity. If the output of the station, either in amperes or watts, is represented by the ordinates of a curve, the abscissae of which represent the hours of the day, this load diagram for a supply station with lighting load only, is a curve such as is shown in fig. 1, having a high peak somewhere between 6 and 8 p.m. The area enclosed by this load-diagram compared with the area of the circumscribing rectangle is called the *load-factor* of the station. This varies from day to day during the year, but on the average for a simple lighting load is not generally above 10 or 12%, and may be lower. Thus the total output from the station is only some 10% on an average of that which it would be if the supply were at all times equal to the maximum demand. Roughly speaking, therefore, the total output of an electric supply station, furnishing current chiefly for electric lighting, is at best equal to about two hours' supply during the day at full load. Hence during the greater part of the twenty-four hours a large part of the plant is lying idle. It is usual to provide certain small sets of steam dynamos, called the daylight machines, for supplying the demand during the day and later part of the evening, the remainder of the machines being called into requisition only for a short time. Provision must be made for sufficient reserve of plant, so that the breakdown of one or more sets will not cripple the output of the station.

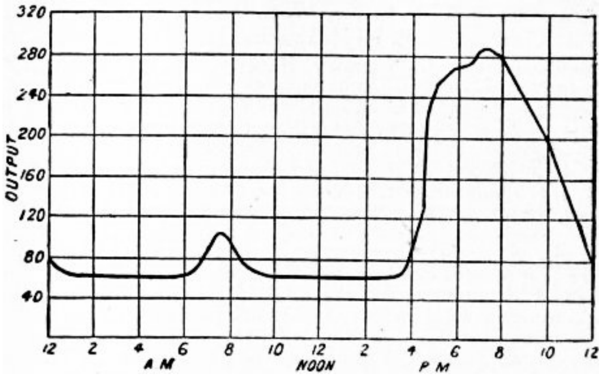


FIG. 1.

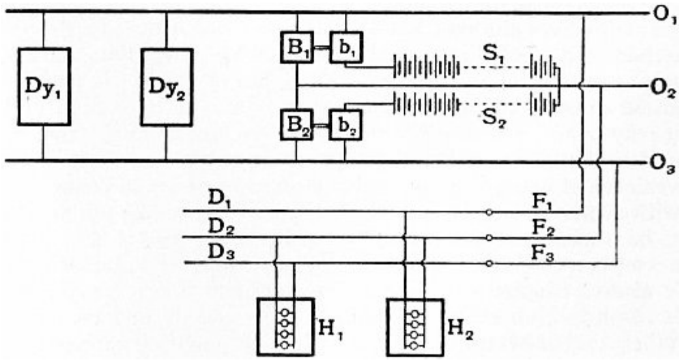
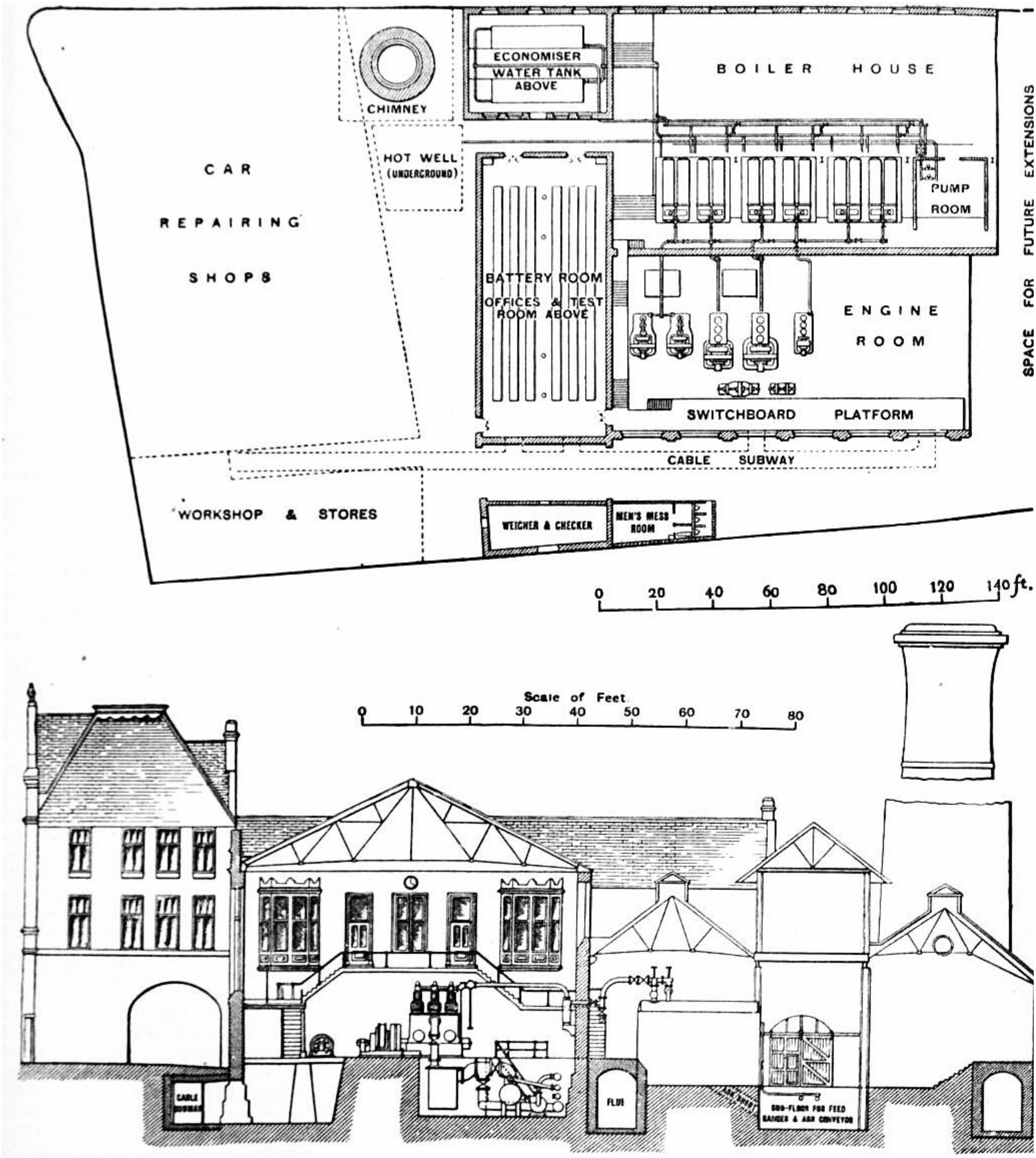


FIG. 2.

Assuming current to be supplied at about 460 volts by different and separate steam dynamos, Dy_1, Dy_2 (fig. 2), the machines are connected through proper amperemeters and voltmeters with *omnibus bars*, O_1, O_2, O_3, on a main switchboard, so that any dynamo can be put in connexion or removed. The switchboard is generally divided into three parts—one panel for the connexions of the positive feeders, F_1, with the positive terminals of the generators; one for the negative feeders, F_3, and negative generator terminals; while from the third (or middle-wire panel) proceed an equal number of middle-wire feeders, F_2. These sets of conductors are led out into the district to be supplied with current, and are there connected into a distributing system, consisting of three separate insulated conductors, D_1, D_2, D_3, respectively called the positive, middle and negative distributing mains. The lamps in the houses, H_1, H_2, &c., are connected between the middle and negative, and the middle and positive, mains by smaller supply and service wires. As far as possible the numbers of lamps installed on the two sides of the system are kept equal; but since it is not possible to control the consumption of current, it becomes necessary to provide at the station two small dynamos called the *balancing machines*, B_1, B_2, connected respectively between the middle and positive and the middle and negative omnibus bars. These machines may have their shafts connected together, or they may be driven by separate steam dynamos; their function is to supply the difference in the total current circulating through the whole of the lamps respectively on the two opposite sides of the middle wire. If storage batteries are employed in the station, it is usual to install two complete batteries, S_1, S_2, which are placed in a separate battery room and connected 195
between the middle omnibus bar and the two outer omnibus bars. The extra electromotive force required to charge these batteries is supplied by two small dynamos b_1, b_2, called *boosters*. It is not unusual to join together the two balancing dynamos and the two boosters on one common bedplate, the shafts being coupled and in line, and to employ the balancing machines as electromotors to drive the boosters as required. By the use of *reversible boosters*, such as those made by the Lancashire Dynamo & Motor Company under the patents of Turnbull & M^c^Leod, having four field windings on the booster magnets (see *The Electrician*, 1904, p. 303), it is possible to adjust the relative duty of the dynamos and battery so that the load on the supply dynamos is always constant. Under these conditions the main engines can be worked all the time at their maximum steam economy and a smaller engine plant employed. If the load in the station rises above the fixed amount, the batteries discharge in parallel with the station dynamos; if it falls below, the batteries are charged and the station dynamos take the external load.

Three-wire system.



From The Electrician.

Figs. 3 and 4.—Low-pressure Supply Station.

The general arrangements of a low-pressure supply station are shown in figs. 3 and 4. It consists of a boiler-house containing a bank of boilers, either Lancashire or Babcock & Wilcox being generally used (see Boiler), which furnish steam to the engines and dynamos, provision being made by duplicate 196

Generating stations.

steam-pipes or a ring main so that the failure of a single engine or dynamo does not cripple the whole supply. The furnace gases are taken through an economizer (generally Green's) so that they give up their heat to the cold feed water. If condensing water is available the engines are worked condensing, and this is an essential condition of economy when steam turbines are employed. Hence, either a condensing water pond or a cooling tower has to be provided to cool the condensing water and enable it to be used over and over again. Preferably the station should be situated near a river or canal and a railway siding. The steam dynamos are generally arranged in an engine-room so as to be overlooked from a switchboard gallery (fig. 3), from which all the control is carried out. The boiler furnaces are usually stoked by automatic stokers. Owing to the relatively small load factor (say 8 or 10%) of a station giving electric supply for lighting only, the object of every station engineer is to cultivate a demand for electric current for power during the day-time by encouraging the use of electric motors for lifts and other purposes, but above all to create a demand for traction purposes. Hence most urban stations now supply current not only for electric lighting but for running the town tramway system, and this traction load being chiefly a daylight load serves to keep the plant employed and remunerative. It is usual to furnish a continuous current supply for traction at 500 or 600 volts, although some station engineers are advocating the use of higher voltages. In those stations which supply current for traction, but which have a widely scattered lighting load, *double current* dynamos are often employed, furnishing from one and the same armature a continuous current for traction purposes, and an alternating current for lighting purposes.

In some places a high voltage system of electric supply by continuous current is adopted. In this case the current is generated at a pressure of 1000 or 2000 volts, and transmitted from the generating station by conductors, called high-pressure feeders, to certain sub-centres or transformer centres, which are either buildings above ground or cellars or excavations under the ground. In these transformer centres are placed machines, called *continuous-current transformers*, which transform the electric energy and create a secondary electric current at a lower pressure, perhaps 100 or 150 volts, to be supplied by distributing mains to users (see TRANSFORMERS). From these sub-centres insulated conductors are run back to the generating station, by which the engineer can start or stop the continuous-current rotatory transformers, and at the same time inform himself as to their proper action and the electromotive force at the secondary terminals. This system was first put in practice in Oxford, England, and hence has been sometimes called by British engineers "the Oxford system." It is now in operation in a number of places in England, such as Wolverhampton, Walsall, and Shoreditch in London. It has the advantage that in connexion with the low-pressure distributing system secondary batteries can be employed, so that a storage of electric energy is effected. Further, continuous-current arc lamps can be worked in series off the high-pressure mains, that is to say, sets of 20 to 40 arc lamps can be operated for the purpose of street lighting by means of the high-pressure continuous current.

High-pressure continuous supply.

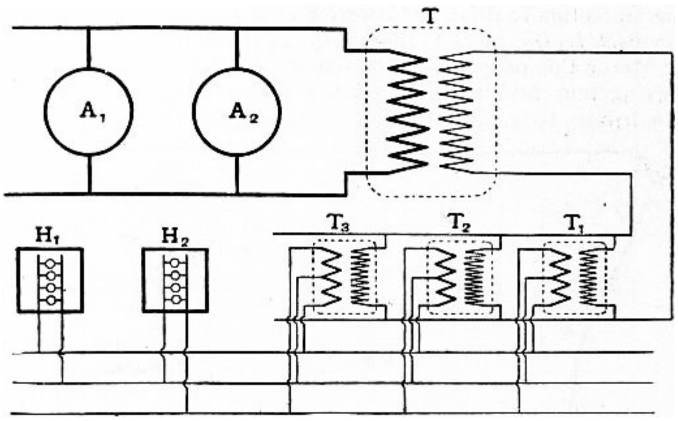


FIG. 5.

The alternating current systems in operation at the present time are the *single-phase* system, with distributing transformers or transformer sub-centres, and the *polyphase* systems, in which the alternating current is transformed down into an alternating current of low pressure, or, by means of rotatory transformers, into a continuous current. The general arrangement of a *single-phase* alternating-current system is as follows: The generating station contains a number of alternators, A_1 A_2 (fig. 5), producing single-phase alternating current, either at 1000, 2000, or sometimes, as at Deptford and other places, 10,000 volts. This current is distributed from the station either at the pressure at which it is generated, or after being transformed up to a higher pressure by the transformer T. The alternators are sometimes worked in parallel, that is to say, all furnish their current to two common omnibus bars on a high-pressure switchboard, and each is switched into circuit at the moment when it is brought into step with the other machines, as shown by some form of *phase-indicator*. In some cases, instead of the high-pressure feeders starting from omnibus bars, each alternator works independently and the feeders are grouped together on the various alternators as required. A number of high-pressure feeders are carried from the main switchboard to various transformer sub-centres or else run throughout the district to which current is to be furnished. If the system laid down is the transformer sub-centre system, then at each of these sub-centres is placed a battery of alternating-current transformers, T_1 T_2 T_3, having their primary circuits all joined in parallel to the terminals of the high-pressure feeders, and their secondary circuits all joined in parallel on a distributing main, suitable switches and cut-outs being interposed. The pressure of the current is then transformed down by these transformers to the required supply pressure. The secondary circuits of these transformers are generally provided with three terminals, so as to supply the low-pressure side on a three-wire system. It is not advisable to connect together directly the secondary circuits of all the different sub-centres, because then a fault or short circuit on one secondary system affects all the others. In banking together transformers in this manner in a sub-station it is necessary to take care that the transformation ratio and secondary drop (see TRANSFORMERS) are exactly the same, otherwise one transformer will take more than its full share of the load and will become overheated. The transformer sub-station system can only be adopted where the area of supply is tolerably compact. Where the consumers lie scattered over a large area, it is necessary to carry the high-pressure mains throughout the area, and to place a separate transformer or transformers in each building. From a financial point of view, this "house-to-house system" of alternating-current supply, generally speaking, is less satisfactory in results than the transformer sub-centre system. In the latter some of the transformers can be switched off, either by hand or by automatic apparatus, during the time when the load is light, and then no power is expended in magnetizing their cores. But with the house-to-house system the whole of the transformers continually remain connected with the high-pressure circuits; hence in the case of supply stations which have only an ordinary electric lighting load, and therefore a load-factor not above 10%, the efficiency of distribution is considerably diminished.

Alternating supply.

The single-phase alternating-current system is defective in that it cannot be readily combined with secondary batteries for the storage of electric energy. Hence in many places preference is now given to the *polyphase system*. In such a system a polyphase alternating current, either two- or three-phase, is transmitted from the generating station at a pressure of 5000 to 10,000 volts, or sometimes higher, and at various sub-stations is transformed down, first by static transformers into an alternating current of lower pressure, say 500 volts, and then by means of rotatory transformers into a continuous current of 500 volts or lower for use for lighting or traction.

In the case of large cities such as London, New York, Chicago, Berlin and Paris the use of small supply stations situated in the interior of the city has gradually given way to the establishment of large supply stations outside the area; in these alternating current is generated on the single or polyphase system at a high voltage and transmitted by underground cables to sub-stations in the city, at which it is transformed down for distribution for private and public electric lighting and for urban electric traction.

Owing to the high relative cost of electric power when generated in small amounts and the great advantages of generating it in proximity to coal mines and waterfalls, the supply of electric power in bulk to small towns and manufacturing districts has become a great feature in modern electrical engineering. In Great Britain, where there is little useful water power but abundance of coal, electric supply stations for supply in bulk have been built in the coal-producing districts of South Wales, the Midlands, the Clyde valley and Yorkshire. In these cases the current is a polyphase current generated at a high voltage, 5000 to 10,000 volts, and sometimes raised again in pressure to 20,000 or 40,000 volts and transmitted by overhead lines to the districts to be supplied. It is there reduced in voltage by transformers and employed as an alternating current, or is used to drive polyphase motors coupled to direct current generators to reproduce the power in continuous current form. It is then distributed for local lighting, street or railway traction, driving motors, and metallurgical or electrochemical applications. Experience has shown that it is quite feasible to distribute in all directions for 25 miles round a high-pressure generating station, which thus supplies an area of nearly 2000 sq. m. At such stations, employing large turbine engines and alternators, electric power may be generated at a works cost of 0.375d. per kilowatt (K.W.), the coal cost being less than 0.125d. per K.W., and the selling price to large load-factor users not more than 0.5d. per K.W. The average price of supply from the local generating stations in towns and cities is from 3d. to 4d. per unit, electric energy for power and heating being charged at a lower rate than that for lighting only.

Conductors.

We have next to consider the structure and the arrangement of the conductors employed to convey the currents from their place of creation to that of utilization. The conductors themselves for the most part consist of copper having a conductivity of not less than 98% according to Matthiessen’s standard. They are distinguished as (1) *External conductors*, which are a part of the public supply and belong to the corporation or company supplying the electricity; (2) *Internal conductors*, or house wiring, forming a part of the structure of the house or building supplied and usually the property of its owner.

External conductors.

The external conductors may be overhead or underground. *Overhead* conductors may consist of bare stranded copper cables carried on porcelain insulators mounted on stout iron or wooden poles. If the current is a high-pressure one, these insulators must be carefully tested, and are preferably of the pattern known as oil insulators. In and near towns it is necessary to employ insulated overhead conductors, generally india-rubber-covered stranded copper cables, suspended by leather loops from steel bearer wires which take the weight. The British Board of Trade have issued elaborate rules for the construction of overhead lines to transmit large electric currents. Where telephone and telegraph wires pass over such overhead electric lighting wires, they have to be protected from falling on the latter by means of guard wires.

By far the largest part, however, of the external electric distribution is now carried out by *underground conductors*, which are either bare or insulated. Bare copper conductors may be carried underground in culverts or chases, air being in this case the insulating material, as in the overhead system. A culvert and covered chase is constructed under the road or side-walk, and properly shaped oak crossbars are placed in it carrying glass or porcelain insulators, on which stranded copper cables, or, preferably, copper strips placed edgeways, are stretched and supported. The advantages of this method of construction are cheapness and the ease with which connexions can be made with service-lines for house supply; the disadvantages are the somewhat large space in which coal-gas leaking out of gas-pipes can accumulate, and the difficulty of keeping the culverts at all times free from rain-water. Moisture has a tendency to collect on the negative insulators, and hence to make a dead earth on the negative side of the main; while unless the culverts are well ventilated, explosions from mixtures of coal-gas and air are liable to occur. Insulated cables are insulated either with a material which is in itself waterproof, or with one which is only waterproof in so far as it is enclosed in a waterproof tube, *e.g.* of lead. Gutta-percha and india-rubber are examples of materials of the former kind. Gutta-percha, although practically everlasting when in darkness and laid under water, as in the case of submarine cables, has not been found satisfactory for use with large systems of electric distribution, although much employed for telephone and telegraph work. Insulated underground external conductors are of three types:—(*a*) *Insulated Cables drawn into Pipes.*—In this system of distribution cast-iron or stoneware pipes, or special stoneware conduits, or conduits made of a material called bitumen concrete, are first laid underground in the street. These contain a number of holes or “ways,” and at intervals drawing-in boxes are placed which consist of a brick or cast-iron box having a water-tight lid, by means of which access is gained to a certain section of the conduit. Wires are used to draw in the cables, which are covered with either india-rubber or lead, the copper being insulated by means of paper, impregnated jute, or other similar material. The advantages of a drawing-in system are that spare ways can be left when the conduits are put in, so that at a future time fresh cables can be added without breaking up the roadway. (*b*) *Cables in Bitumen.*—One of the earliest systems of distribution employed by T.A. Edison consisted in fixing two segment-shaped copper conductors in a steel tube, the interspace between the conductors and the tube being filled in with a bitumen compound. A later plan is to lay down an iron trough, in which the cables are supported by wooden bearers at proper distances, and fill in the whole with natural bitumen. This system has been carried out extensively by the Callendar Cable Company. Occasionally concentric lead-covered and armoured cables are laid in this way, and then form an expensive but highly efficient form of insulated conductor. In selecting a system of distribution regard must be paid to the nature of the soil in which the cables are laid. Lead is easily attacked by soft water, although under some conditions it is apparently exceedingly durable, and an atmosphere containing coal-gas is injurious to india-rubber. (*c*) *Armoured Cables.*—In a very extensively used system of distribution armoured cables are employed. In this case the copper conductors, two, three or more in number, may be twisted together or arranged concentrically, and insulated by means of specially prepared jute or paper insulation, overlaid with a continuous tube of lead.

Over the lead, but separated by a hemp covering, is put a steel armour consisting of two layers of steel strip, wound in opposite directions and kept in place by an external covering. Such a cable can be laid directly in the ground without any preparation other than the excavation of a simple trench, junction-boxes being inserted at intervals to allow of branch cables being taken off. The armoured cable used is generally of the concentric pattern (fig. 6). It consists of a stranded copper cable composed of a number of wires twisted together and overlaid with an insulating material. Outside this a tubular arrangement of copper wires and a second layer of insulation, and finally a protective covering of lead and steel wires or armour are placed. In some cases three concentric cylindrical conductors are formed by twisting wires or copper strips with insulating material between. In others two or three cables of stranded copper are embedded in insulating material and included in a lead sheath. This last type of cable is usually called a *two-* or *three-core* pattern cable (fig. 7).

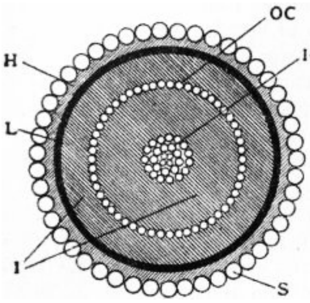


Fig. 6.—Armoured Concentric Cable (Section).

IC, Inner conductor.
OC, Outer conductor.
I, Insulation.
L, Lead sheath.
S, Steel armour.
H, Hemp covering.

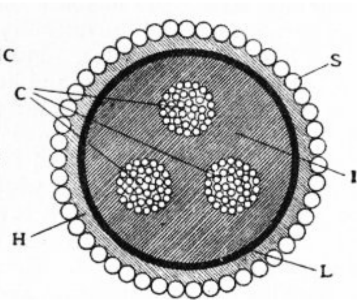


Fig. 7.—Triple Conductor Armoured Cable (Section).

C, Copper conductor.
I, Insulation.
L, Lead sheath.
H, Hemp covering.
S, Steel armour.

The arrangement and nature of the external conductors depends on the system of electric supply in which they are used. In the case of continuous-current supply for incandescent electric lighting and motive power in small units, when the external conductors are laid down on the three-wire system, each main or branch cable in the street consists of a set of three conductors called the positive, middle and negative. Of these triple conductors some run from the supply station to various points in the area of supply without being tapped, and are called the *feeders*; others, called the *distributing mains*, are used for making connexions with the service lines of the consumers, one service line, as already explained, being connected to the middle conductor, and the other to either the positive or the negative one. Since the middle conductor serves to convey only the difference between the currents being used on the two sides of the system, it is smaller in section than the positive and negative ones. In laying out the system great judgment has to be exercised as to the selection of the points of attachment of the feeders to the distributing mains, the object being to keep a constant electric pressure or voltage between the two service-lines in all the houses independently of the varying demand for current. Legally the suppliers are under regulations to keep the supply voltage constant within 4% either way above or below the standard pressure. As a matter of fact very few stations do maintain such good regulation. Hence a considerable variation in the light given by the incandescent lamps is observed, since the candle-power of carbon glow lamps varies as the fifth or sixth power of the voltage of supply, *i.e.* a variation of only 2% in the supply pressure affects the resulting candle-power of the lamps to the extent of 10 or 12%. This variation is, however, less in the case of metallic filament lamps (see Lighting: *Electric*). In the service-lines are inserted the meters for measuring the electric energy supplied to the customer (see Meter, Electric).

Interior wiring.

In the interior of houses and buildings the conductors generally consist of india-rubber-covered cables laid in wood casing. The copper wire must be tinned and then covered, first with a layer of unvulcanized pure india-rubber, then with a layer of vulcanized rubber, and lastly with one or more layers of protective cotton twist or tape. No conductor of this character employed for interior house-wiring should have a smaller insulation resistance than 300 megohms per mile when tested with a pressure of 600 volts after soaking 24 hours in water. The wood casing should, if placed in damp positions or under plaster, be well varnished with waterproof varnish. As far as possible all joints in the run of the cable should be avoided by the use of the so-called looping-in system, and after the wiring is complete, careful tests for insulation should be made. The Institution of Electrical Engineers of Great Britain have drawn up rules to be followed in interior house-wiring, and the principal Fire Insurance offices, following the lead of the Phoenix Fire Office, of London, have made regulations which, if followed, are a safeguard against bad workmanship and resulting possibility of damage by fire. Where fires having an electric origin have taken place, they have invariably been traced to some breach of these rules. Opinions differ, however, as to the value and security of this method of laying interior conductors in buildings, and two or three alternative systems have been much employed. In one of these, called the *interior conduit* system, highly insulating waterproof and practically fireproof tubes or conduits replace the wooden casing; these, being either of plain insulating material, or covered with brass or steel armour, may be placed under plaster or against walls. They are connected by bends or joint-boxes. The insulated wires being drawn into them, any short circuit or heating of the wire cannot give rise to a fire, as it can only take place in the interior of a non-inflammable tube. A third system of electric light wiring is the safety concentric system, in which concentric conductors are used. The inner one, which is well insulated, consists of a copper-stranded cable. The outer may be a galvanized iron strand, a copper tape or braid, or a brass tube, and is therefore necessarily connected with the earth. A fourth system consists in the employment of twin insulated wires twisted together and sheathed with a lead tube; the conductor thus formed can be fastened by staples against walls, or laid under plaster or floors.

The general arrangement for distributing current to the different portions of a building for the purpose of electric lighting is to run up one or more rising mains, from which branches are taken off to distributing boxes on each floor, and from these boxes to carry various branch circuits to the lamps. At the distributing boxes are collected the cut-outs and switches controlling the various circuits. When alternating currents are employed, it is usual to select as a type of conductor either twin-twisted conductor or concentric; and the employment of these types of cable, rather than two separate cables, is essential in any case where there are telephone or telegraph wires in proximity, for otherwise the alternating current would create inductive disturbances in the telephone circuit. The house-wiring also comprises the details of *switches* for controlling the lamps, *cut-outs* or fuses for preventing an excess of current passing, and fixtures or supports for lamps often of an ornamental character. For the details of these, special treatises on electric interior wiring must be consulted.

For further information the reader may be referred to the following books:—C.H. Wordingham, *Central Electrical Stations* (London, 1901); A. Gay and C.Y. Yeaman, *Central Station Electricity Supply* (London, 1906); S.P. Thompson, *Dynamo Electric Machinery* (2 vols., London, 1905); E. Tremlett Carter and T. Davies, *Motive Power and Gearing* (London, 1906); W.C. Clinton, *Electric Wiring* (2nd ed., London, 1906); W. Perren Maycock, *Electric Wiring, Fitting, Switches and Lamps* (London, 1899); D. Salomons, *Electric Light Installations* (London, 1894); Stuart A. Russell, *Electric Light Cables* (London, 1901); F.A.C. Perrine, *Conductors for Electrical Distribution* (London, 1903); E. Rosenberg, W.W. Haldane Gee and C. Kinzbrunner, *Electrical Engineering* (London, 1903); E.C. Metcalfe, *Practical Electric Wiring for Lighting Installations* (London, 1905); F.C. Raphael, *The Wireman’s Pocket Book* (London, 1903).

(J. A. F.)

History.

II. *Commercial Aspects.*—To enable the public supply enterprises referred to in the foregoing section to be carried out in England, statutory powers became necessary to break up the streets. In the early days a few small stations were established for the supply of electricity within “block” buildings, or by means of overhead wires within restricted areas, but the limitations proved uneconomical and the installations were for the most part merged into larger undertakings sanctioned by parliamentary powers. In the year 1879 the British government had its attention directed for the first time to electric lighting as a possible subject for legislation, and the consideration of the then existing state of electric lighting was referred to a select committee of the House of Commons. No legislative action, however, was taken at that time. In fact the invention of the incandescent lamp was incomplete—Edison’s British master-patent was only filed in Great Britain in November 1879. In 1881 and 1882 electrical exhibitions were held in Paris and at the Crystal Palace, London, where the improved electric incandescent lamp was brought 199
before the general public. In 1882 parliament passed the first Electric Lighting Act, and considerable speculation ensued. The aggregate capital of the companies registered in 1882-1883 to carry out the public supply of electricity in the United Kingdom amounted to £15,000,000, but the onerous conditions of the act deterred investors from proceeding with the enterprise. Not one of the sixty-two provisional orders granted to companies in 1883 under the act was carried out. In 1884 the Board of Trade received only four applications for provisional orders, and during the subsequent four years only one order was granted. Capitalists declined to go on with a business which if successful could be taken away from them by local authorities at the end of twenty-one years upon terms of paying only the then value of the plant, lands and buildings, without regard to past or future profits, goodwill or other considerations. The electrical industry in Great Britain ripened at a time when public opinion was averse to the creation of further monopolies, the general belief being that railway, water and gas companies had in the past received valuable concessions on terms which did not sufficiently safeguard the interests of the community. The great development of industries by means of private enterprise in the early part of the 19th century produced a reaction which in the latter part of the century had the effect of discouraging the creation by private enterprise of undertakings partaking of the nature of monopolies; and at the same time efforts were made to strengthen local and municipal institutions by investing them with wider functions. There were no fixed principles governing the relations between the state or municipal authorities and commercial companies rendering monopoly services. The new conditions imposed on private enterprise for the purpose of safeguarding the interests of the public were very tentative, and a former permanent secretary of the Board of Trade has stated that the efforts made by parliament in these directions have sometimes proved injurious alike to the public and to investors. One of these tentative measures was the Tramways Act 1870, and twelve years later it was followed by the first Electric Lighting Act.

It was several years before parliament recognized the harm that had been done by the passing of the Electric Lighting Act 1882. A select committee of the House of Lords sat in 1886 to consider the question of reform, and as a result the Electric Lighting Act 1888 was passed. This amending act altered the period of purchase from twenty-one to forty-two years, but the terms of purchase were not materially altered in favour of investors. The act, while stipulating for the consent of local authorities to the granting of provisional orders, gives the Board of Trade power in exceptional cases to dispense with the consent, but this power has been used very sparingly. The right of vetoing an undertaking, conferred on local authorities by the Electric Lighting Acts and also by the Tramways Act 1870, has frequently been made use of to exact unduly onerous conditions from promoters, and has been the subject of complaint for years. Although, in the opinion of ministers of the Crown, the exercise of the veto by local authorities has on several occasions led to considerable scandals, no government has so far been able, owing to the very great power possessed by local authorities, to modify the law in this respect. After 1888 electric lighting went ahead in Great Britain for the first time, although other countries where legislation was different had long previously enjoyed its benefits. The developments proceeded along three well-defined lines. In London, where none of the gas undertakings was in the hands of local authorities, many of the districts were allotted to companies, and competition was permitted between two and sometimes three companies. In the provinces the cities and larger towns were held by the municipalities, while the smaller towns, in cases where consents could be obtained, were left to the enterprise of companies. Where consents could not be obtained these towns were for some time left without supply.

Some statistics showing the position of the electricity supply business respectively in 1896 and 1906 are interesting as indicating the progress made and as a means of comparison between these two periods of the state of the industry as a whole. In 1896 thirty-eight companies were at work with an aggregate capital of about £6,000,000, and thirty-three municipalities with electric lighting loans of nearly £2,000,000. The figures for 1906, ten years later, show that 187 electricity supply companies were in operation with a total investment of close on £32,000,000, and 277 municipalities with loans amounting to close on £36,000,000. The average return on the capital invested in the companies at the later period was 5.1% per annum. In 1896 the average capital expenditure was about £100 per kilowatt of plant installed; and £50 per kilowatt was regarded as a very low record. For 1906 the average capital expenditure per kilowatt installed was about £81. The main divisions of the average expenditure are:—

	1896.	1906.
Land and buildings	22.3%	17.8%
Plant and machinery	36.7	36.5
Mains	32.2	35.5
Meters and instruments	4.6	5.7
Provisional orders, &c.	3.2	2.8

The load connected, expressed in equivalents of eight candle-power lamps, was 2,000,000 in 1896 and 24,000,000 in 1906. About one-third of this load would be for power purposes and about two-thirds for lighting. The Board of Trade units sold were 30,200,000 in 1896 and 533,600,000 in 1906, and the average prices per unit obtained were 5.7d. and 2.7d. respectively, or a revenue of £717,250 in 1896 and over £6,000,000 in 1906. The working expenses per Board of Trade unit sold, excluding depreciation, sinking fund and interest were as follows:—

	1896.	1906.
Generation and distribution	2.81d.	.99d.
Rent, rates and taxes	.35	.14
Management	.81	.18
Sundries	.10	.02
	———	———
Total	4.07d.	1.33d.

In 1896 the greatest output at one station was about 5½ million units, while in 1906 the station at Manchester had the largest output of over 40 million units.

The capacity of the plants installed in the United Kingdom in 1906 was:—

	K.W.		
Continuous current	417,000	Provinces	333,000
		London	84,000
Alternating current	132,000	Provinces	83,000
		London	49,000
Continuous current and alternating current combined	480,000	Provinces	366,000
		London	114,000
	————		
	1,029,000	k.w.	

Economics.

The economics of electric lighting were at first assumed to be similar to those of gas lighting. Experience, however, soon proved that there were important differences, one being that gas may be stored in gasometers without appreciable loss and the work of production carried on steadily without reference to fluctuations of demand. Electricity cannot be economically stored to the same extent, and for the most part it has to be used as it is generated. The demand for electric light is practically confined to the hours between sunset and midnight, and it rises sharply to a "peak" during this period. Consequently the generating station has to be equipped with plant of sufficient capacity to cope with the maximum load, although the peak does not persist for many minutes—a condition which is very uneconomical both as regards capital expenditure and working costs (see Lighting: *Electric*). In order to obviate the unproductiveness of the generating plant during the greater part of the day, electricity supply undertakings sought to develop the "daylight" load. This they did by supplying electricity for traction purposes, but more particularly for industrial power purposes. The difficulties in the way of this line of development, however, were that electric power could not be supplied cheaply enough to compete with steam, hydraulic, gas and other forms of power, unless it was generated on a very large scale, and this large demand could not be developed within the restricted areas for which provisional orders were granted and under the restrictive conditions of these orders in regard to situation of power-house and other matters.

The leading factors which make for economy in electricity supply are the magnitude of the output, the load factor, and the diversity factor, also the situation of the power house, the means of distribution, and the 200
provision of suitable, trustworthy and efficient plant. These factors become more favourable the larger the area and the greater and more varied the demand to be supplied. Generally speaking, as the output increases so the cost per unit diminishes, but the ratio (called the load factor) which the output during any given period bears to the *maximum* possible output during the same period has a very important influence on costs. The

ideal condition would be when a power station is working at its normal *maximum* output continuously night and day. This would give a load-factor of 100%, and represents the ultimate ideal towards which the electrical engineer strives by increasing the area of his operations and consequently also the load and the variety of the overlapping demands. It is only by combining a large number of demands which fluctuate at different times—that is by achieving a high diversity factor—that the supplier of electricity can hope to approach the ideal of continuous and steady output. Owing to the dovetailing of miscellaneous demands the actual demand on a power station at any moment is never anything like the aggregate of all the maximum demands. One large station would require a plant of 36,000 k.w. capacity if all the demands came upon the station simultaneously, but the maximum demand on the generating plant is only 15,000 kilowatts. The difference between these two figures may be taken to represent the economy effected by combining a large number of demands on one station. In short, the keynote of progress in cheap electricity is increased and diversified demand combined with concentration of load. The average load-factor of all the British electricity stations in 1907 was 14.5%—a figure which tends to improve.

Power companies.

Several electric power supply companies have been established in the United Kingdom to give practical effect to these principles. The Electric Lighting Acts, however, do not provide for the establishment of large power companies, and special acts of parliament have had to be promoted to authorize these undertakings. In 1898 several bills were introduced in parliament for these purposes. They were referred to a joint committee of both Houses of Parliament presided over by Lord Cross. The committee concluded that, where sufficient public advantages are shown, powers should be given for the supply of electricity over areas including the districts of several local authorities and involving the use of exceptional plant; that the usual conditions of purchase of the undertakings by the local authorities did not apply to such undertakings; that the period of forty-two years was “none too long” a tenure; and that the terms of purchase should be reconsidered. With regard to the provision of the Electric Lighting Acts which requires that the consent of the local authority should be obtained as a condition precedent to the granting of a provisional order, the committee was of opinion that the local authority should be entitled to be heard by the Board of Trade, but should not have the power of veto. No general legislation took place as a result of these recommendations, but the undermentioned special acts constituting power supply companies were passed.

In 1902 the president of the Board of Trade stated that a bill had been drafted which he thought “would go far to meet all the reasonable objections that had been urged against the present powers by the local authorities.” In 1904 the government introduced the Supply of Electricity Bill, which provided for the removal of some of the minor anomalies in the law relating to electricity. The bill passed through all its stages in the House of Lords but was not proceeded with in the House of Commons. In 1905 the bill was again presented to parliament but allowed to lie on the table. In the words of the president of the Board of Trade, there was “difficulty of dealing with this question so long as local authorities took so strong a view as to the power which ought to be reserved to them in connexion with this enterprise.” In the official language of the council of the Institution of Electrical Engineers, the development of electrical science in the United Kingdom is in a backward condition as compared with other countries in respect of the practical application to the industrial and social requirements of the nation, notwithstanding that Englishmen have been among the first in inventive genius. The cause of such backwardness is largely due to the conditions under which the electrical industry has been carried on in the country, and especially to the restrictive character of the legislation governing the initiation and development of electrical power and traction undertakings, and to the powers of obstruction granted to local authorities. Eventually The Electric Lighting Act 1909 was passed. This Act provides:—(1) for the granting of provisional orders authorizing any local authority or company to supply electricity in bulk; (2) for the exercise of electric lighting powers by local authorities jointly under provisional order; (3) for the supply of electricity to railways, canals and tramways outside the area of supply with the consent of the Board of Trade; (4) for the compulsory acquisition of land for generating stations by provisional order; (5) for the exemption of agreements for the supply of electricity from stamp duty; and (6) for the amendment of regulations relating to July notices, revision of maximum price, certification of meters, transfer of powers of undertakers, auditors’ reports, and other matters.

The first of the Power Bills was promoted in 1898, under which it was proposed to erect a large generating station in the Midlands from which an area of about two thousand square miles would be supplied. Vigorous opposition was organized against the bill by the local authorities and it did not pass. The bill was revived in 1899, but was finally crushed. In 1900 and following years several power bills were successfully promoted, and the following are the areas over which the powers of these acts extend:

In Scotland, (1) the Clyde Valley, (2) the county of Fife, (3) the districts described as “Scottish Central,” comprising Linlithgow, Clackmannan, and portions of Dumbarton and Stirling, and (4) the Lothians, which include portions of Midlothian, East Lothian, Peebles and Lanark.

In England there are companies operating in (1) Northumberland, (2) Durham county, (3) Lancashire, (4) South Wales and Carmarthenshire, (5) Derbyshire and Nottinghamshire, (6) Leicestershire and Warwickshire, (7) Yorkshire, (8) Shropshire, Worcestershire and Staffordshire, (9) Somerset, (10) Kent, (11) Cornwall, (12) portions of Gloucestershire, (13) North Wales, (14) North Staffordshire, Derbyshire, Denbighshire and Flintshire, (15) West Cumberland, (16) the Cleveland district, (17) the North Metropolitan district, and (18) the West Metropolitan area. An undertaking which may be included in this category, although it is not a Power Act company, is the Midland Electric Corporation in South Staffordshire. The systems of generation and distribution are generally 10,000 or 11,000 volts three-phase alternating current.

The powers conferred by these acts were much restricted as a result of opposition offered to them. In many cases the larger towns were cut out of the areas of supply altogether, but the general rule was that the power company was prohibited from supplying direct to a power consumer in the area of an authorized distributor without the consent of the latter, subject to appeal to the Board of Trade. Even this restricted power of direct supply was not embodied in all the acts, the power of taking supply in bulk being left only to certain authorized distributors and to authorized users such as railways and tramways. Owing chiefly to the exclusion of large towns and industrial centres from their areas, these power supply companies did not all prove as successful as was expected.

In the case of one of the power companies which has been in a favourable position for the development of its business, the theoretical conclusions in regard to the economy of large production above stated have been amply demonstrated in practice. In 1901, when this company was emerging from the stage of a simple electric lighting company, the total costs per unit were 1.05d. with an output of about 2½ million units per annum. In 1905 the output rose to over 30 million units mostly for power and traction purposes, and the costs fell to 0.56d. per unit.

An interesting phase of the power supply question has arisen in London. Under the general acts it was stipulated that the power-house should be erected within the area of supply, and amalgamation of 201
undertakings was prohibited. After less than a decade of development several of the companies in London found themselves obliged to make considerable additions to their generating plants. But their existing buildings were full to their utmost capacity, and the difficulties of generating cheaply on crowded sites had increased instead of diminished during the interval. Several of the companies had to promote special acts of parliament to obtain relief, but the idea of a general combination was not considered to be within the range of practical politics until 1905, when the Administrative County of London Electric Power Bill was introduced. Compared with other large cities, the consumption of electricity in London is small. The output of electricity in New York for all purposes is 971 million units per annum or 282 units per head of population. The output of electricity in London is only 42 units per head per annum. There are in London twelve local authorities and fourteen companies carrying on electricity supply undertakings. The capital expenditure is £3,127,000 by the local authorities and £12,530,000 by the companies, and their aggregate capacity of plant is 165,000 k.w. The total output is about 160,000,000 units per annum, the total revenue is over £2,000,000, and the gross profit before providing for interest and sinking fund charges is £1,158,000. The general average cost of production is 1.55d. per unit, and the average price per unit sold is 3.16d., but some of the undertakers have already supplied electricity to large power consumers at below 1d. per unit. By generating on a large scale for a wide variety of demands the promoters of the new scheme calculated to be able to offer electrical energy in bulk to electricity supply companies and local authorities at prices substantially below their costs of production at separate stations, and also to provide them and power users with electricity at rates which would compete with other forms of power. The authorized capital was fixed at £6,666,000, and the initial outlay on the first plant of 90,000 k.w., mains, &c., was estimated at £2,000,000. The costs of generation were estimated at 0.15d. per unit, and the total cost at 0.52d. per unit sold. The output by the year 1911 was estimated at 133,500,000 units at an average selling price of 0.7d. per unit, to be reduced to 0.55d. by 1916 when the output was estimated at 600,000,000 units. The bill underwent a searching examination before the House of Lords committee and was passed in an amended form. At the second reading in the House of Commons a strong effort was made to throw it out, but it was allowed to go to committee on the condition—contrary to the general recommendations of the parliamentary committee of 1898—that a purchase clause would be inserted; but amendments were proposed to such an extent that the bill was not reported for third reading until the eve of the prorogation of parliament. In the following year (1906) the Administrative Company’s bill was again introduced in parliament, but the London County Council, which had previously adopted an attitude both hostile and negative, also brought forward a similar bill. Among other schemes, one known as the Additional Electric Power Supply Bill was to authorize the transmission of current from St Neots in Hunts. This bill was rejected by the House of Commons because the promoters declined to give precedence to the bill of the London County Council. The latter bill was referred to a hybrid committee with instructions to consider the whole question of London power supply, but it was ultimately rejected. The same result attended a second bill which was promoted by the London County Council in 1907. The question was settled by the London Electric Supply Act 1908, which constitutes the London County Council the purchasing authority (in the place of the local authorities) for the electric supply companies in London. This Act also enabled the Companies and other authorized undertakers to enter into agreements for the exchange of current and the linking-up of stations.

Legislation and regulations.

The general supply of electricity is governed primarily by the two acts of parliament passed in 1882 and 1888, which apply to the whole of the United Kingdom. Until 1899 the other statutory provisions relating to electricity supply were incorporated in provisional orders granted by the Board of Trade and confirmed by parliament in respect of each undertaking, but in that year an Electric Lighting Clauses Act was passed by which the clauses previously inserted in each order were standardized. Under these acts the Board of Trade made rules with respect to applications for licences and provisional orders, and regulations for the protection of the public, and of the electric lines and works of the post office, and others, and also drew up a model form for provisional orders.

Until the passing of the Electric Lighting Acts, wires could be placed wherever permission for doing so could be obtained, but persons breaking up streets even with the consent of the local authority were liable to indictment for nuisance. With regard to overhead wires crossing the streets, the local authorities had no greater power than any member of the public, but a road authority having power to make a contract for lighting the road could authorize others to erect poles and wires for the purpose. A property owner, however, was able to prevent wires from being taken over his property. The act of 1888 made all electric lines or other works for the supply of electricity, not entirely enclosed within buildings or premises in the same occupation, subject to regulations of the Board of Trade. The postmaster-general may also impose conditions for the protection of the post office. Urban authorities, the London County Council, and some other corporations have now powers to make by-laws for prevention of obstruction from posts and overhead wires for telegraph, telephone, lighting or signalling purposes; and electric lighting stations are now subject to the provisions of the Factory Acts.

Parliamentary powers to supply electricity can now be obtained by (A) Special Act, (B) Licence, or (C) Provisional order.

A. *Special Act.*—Prior to the report of Lord Cross’s joint committee of 1898 (referred to above), only one special act was passed. The provisions of the Electric Power Acts passed subsequently are not uniform, but the following are some of the usual provisions:—

The company shall not supply electricity for lighting purposes except to authorized undertakers, provided that the energy supplied to any person for power may be used for lighting any premises on which the power is utilized. The company shall not supply energy (except to authorized undertakers) in any area which forms part of the area of supply of any authorized distributors without their consent, such consent not to be unreasonably withheld. The company is bound to supply authorized undertakers upon receiving notice and upon the applicants agreeing to pay for at least seven years an amount sufficient to yield 20% on the outlay (excluding generating plant or wires already installed). Other persons to whom the company is authorized to supply may require it upon terms to be settled, if not agreed, by the Board of Trade. Dividends are usually restricted to 8%, with a provision that the rate may be increased upon the average price charged being reduced. The maximum charges are usually limited to 3d. per unit for any quantity up to 400 hours’ supply, and 2d. per unit beyond. No preference is to be shown between consumers in like circumstances. Many provisions of the general Electric Lighting Acts are excluded from these special acts, in particular the clause giving the local authority the right to purchase the undertaking compulsorily.

B. *Licence.*—The only advantages of proceeding by licence are that it can be expeditiously obtained and does not require confirmation by parliament; but some of the provisions usually inserted in provisional orders would be *ultra vires* in a licence, and the Electric Lighting Clauses Act 1899 does not extend to licences. The term of a licence does not exceed seven years, but is renewable. The consent of the local authority is necessary even to an application for a licence. None of the licences that have been granted is now in force.

C. *Provisional Order.*—An intending applicant for a provisional order must serve notice of his intention on every local authority within the proposed area of supply on or before the 1st of July prior to the session in which application is to be made to the Board of Trade. This provision has given rise to much complaint, as it gives the local authorities a long time for bargaining and enables them to supersede the company’s 202
application by themselves applying for provisional orders. The Board of Trade generally give preference to the applications of local authorities.

In 1905 the Board of Trade issued a memorandum stating that, in view of the revocation of a large number of provisional orders which had been obtained by local authorities, or in regard to which local authorities had entered into agreements with companies for carrying the orders into effect (which agreements were in many cases *ultra vires* or at least of doubtful validity), it appeared undesirable that a local authority should apply for a provisional order without having a definite intention of exercising the powers, and that in future the Board of Trade would not grant an order to a local authority unless the board were satisfied that the powers would be exercised within a specified period.

Every undertaking authorized by provisional order is subject to the provision of the general act entitling the local authority to purchase compulsorily at the end of forty-two years (or shorter period), or after the expiration of every subsequent period of ten years (unless varied by agreement between the parties with the consent of the Board of Trade), so much of the undertaking as is within the jurisdiction of the purchasing authority upon the terms of paying the then value of all lands, buildings, works, materials and plant, suitable to and used for the purposes of the undertaking; provided that the value of such lands, &c., shall be deemed to be their fair market value at the time of purchase, due regard being had to the nature and then condition and state of repair thereof, and to the circumstance that they are in such positions as to be ready for immediate working, and to the suitability of the same to the purposes of the undertaking, and where a part only of the undertaking is purchased, to any loss occasioned by severance, but without any addition in respect of compulsory purchase or of goodwill, or of any profits which may or might have been or be made from the undertaking or any similar consideration. Subject to this right of purchase by the local authority, a provisional order (but not a licence) may be for such period as the Board of Trade may think proper, but so far no limit has been imposed, and unless purchased by a local authority the powers are held in perpetuity. No monopoly is granted to undertakers, and since 1889 the policy of the Board of Trade has been to sanction two undertakings in the same metropolitan area, preferably using different systems, but to discourage competing schemes within the same area in the provinces. Undertakers must within two years lay mains in certain specified streets. After the first eighteen months they may be required to lay mains in other streets upon conditions specified in the order, and any owner or occupier of premises within 50 yds. of a distributing main may require the undertakers to give a supply to his premises; but the consumer must pay the cost of the lines laid upon his property and of so much outside as exceeds 60 ft. from the main, and he must also contract for two and in some cases for three years’ supply. But undertakers are prohibited in making agreements for supply from showing any undue preference. The maximum price in London is 13s. 4d. per quarter for any quantity up to 20 units, and beyond that 8d. per unit, but 11s. 8d. per quarter up to 20 units and 7d. per unit beyond is the more general maximum. The “Bermondsey clause” requires the undertakers (local authority) so to fix their charges (not exceeding the specified maximum) that the revenue shall not be less than the expenditure.

There is no statutory obligation on municipalities to provide for depreciation of electricity supply undertakings, but after providing for all expenses, interest on loans, and sinking fund instalments, the local authority may create a reserve fund until it amounts, with interest, to one-tenth of the aggregate capital expenditure. Any deficiency when not met out of reserve is payable out of the local rates.

The principle on which the Local Government Board sanctions municipal loans for electric lighting undertakings is that the period of the loan shall not exceed the life of the works, and that future ratepayers shall not be unduly burdened. The periods of the loans vary from ten years for accumulators and arc lamps to sixty years for lands. Within the county of London the loans raised by the metropolitan borough councils for electrical purposes are sanctioned by the London County Council, and that body allows a minimum period of twenty years for repayment. Up to 1904-1905, 245 loans had been granted by the council amounting in the aggregate to £4,045,067.

In 1901 the Institution of Civil Engineers appointed a committee to consider the advisability of standardizing various kinds of iron and steel sections. Subsequently the original reference was enlarged, and

Standardization. in 1902 the Institution of Electrical Engineers was invited to co-operate. The treasury, as well as railway companies, manufacturers and others, have made grants to defray the expenses. The committee on electrical plant has ten sub-committees. In August 1904 an interim report was issued by the sub-committee on generators, motors and transformers, dealing with pressures and frequencies, rating of generators and motors, direct-current generators, alternating-current generators, and motors.

In 1903 the specification for British standard tramway rails and fish-plates was issued, and in 1904 a standard specification for tubular tramway poles was issued. A sectional committee was formed in 1904 to correspond with foreign countries with regard to the formation of an electrical international commission to study the question of an international standardization of nomenclature and ratings of electrical apparatus and machinery.

The electrical industry. The electrical manufacturing branch, which is closely related to the electricity supply and other operating departments of the electrical industry, only dates from about 1880. Since that time it has undergone many vicissitudes. It began with the manufacture of small arc lighting equipments for railway stations, streets and public buildings. When the incandescent lamp became a commercial article, ship-lighting sets and installations for theatres and mansions constituted the major portion of the electrical work. The next step was the organization of house-to-house distribution of electricity from small "central stations," ultimately leading to the comprehensive public supply in large towns, which involved the manufacture of generating and distributing plants of considerable magnitude and complexity. With the advent of electric traction about 1896, special machinery had to be produced, and at a later stage the manufacturer had to solve problems in connexion with bulk supply in large areas and for power purposes. Each of these main departments involved changes in ancillary manufactures, such as cables, switches, transformers, meters, &c., so that the electrical manufacturing industry has been in a constant state of transition. At the beginning of the period referred to Germany and America were following the lead of England in theoretical developments, and for some time Germany obtained electrical machinery from England. Now scarcely any electrical apparatus is exported to Germany, and considerable imports are received by England from that country and America. The explanation is to be found mainly in the fact that the adverse legislation of 1882 had the effect of restricting enterprise, and while British manufacturers were compulsorily inert during periods of impeded growth of the two most important branches of the industry—electric lighting and traction—manufacturers in America and on the continent of Europe, who were in many ways encouraged by their governments, devoted their resources to the establishment of factories and electrical undertakings, and to the development of efficient selling organizations at home and abroad. When after the amendment of the adverse legislation in 1888 a demand for electrical machinery arose in England, the foreign manufacturers were fully organized for trade on a large scale, and were further aided by fiscal conditions to undersell English manufacturers, not only in neutral markets, but even in their own country. Successful manufacture on a large scale is possible only by standardizing the methods of production. English manufacturers were not able to standardize because they had not the necessary output. There had been no repetitive demand, and there was no production on a large scale. Foreign manufacturers, however, were able to standardize by reason of the large uniform demand which existed for their manufactures. Statistics are 203
available showing the extent to which the growth of the electrical manufacturing industry in Great Britain was delayed. Nearly twenty years after the inception of the industry there were only twenty-four manufacturing companies registered in the United Kingdom, having an aggregate subscribed capital of under £7,000,000. But in 1907 there were 292 companies with over £42,000,000 subscribed capital. The cable and incandescent lamp sections show that when the British manufacturers are allowed opportunities they are not slow to take advantage of them. The cable-making branch was established under the more encouraging conditions of the telegraph industry, and the lamp industry was in the early days protected by patents. Other departments not susceptible to foreign competition on account of freightage, such as the manufacture of storage batteries and rolling stock, are also fairly prosperous. In departments where special circumstances offer a prospect of success, the technical skill, commercial enterprise and general efficiency of British manufacturers manifest themselves by positive progress and not merely by the continuance of a struggle against adverse conditions. The normal posture of the British manufacturer of electrical machinery has been described as one of desperate defence of his home trade; that of the foreign manufacturer as one of vigorous attack upon British and other open markets. In considering the position of English manufacturers as compared with their foreign rivals, some regard should be had to the patent laws. One condition of a grant of a patent in most foreign countries is that the patent shall be worked in those countries within a specified period. But a foreign inventor was until 1907 able to secure patent protection in Great Britain without any obligation to manufacture there. The effect of this was to encourage the manufacture of patented apparatus in foreign countries, and to stimulate their exportation to Great Britain in competition with British products. With regard to the electrochemical industry the progress which has been achieved by other nations, notably Germany, is very marvellous by comparison with the advance made by England, but to state the reasons why this industry has had such extraordinary development in Germany, notwithstanding that many of the fundamental inventions were made in England, would require a statement of the marked differences in the methods by which industrial progress is promoted in the two countries.

There has been very little solidarity among those interested in the commercial development of electricity, and except for the discussion of scientific subjects there has been very little organization with the object of protecting and promoting common interests.

(E. Ga.)

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1 British Patent Specification, No. 5306 of 1878, and No. 602 of 1880.

2 *Ibid.* No. 3988 of 1878.

ELECTRIC WAVES. § 1. Clerk Maxwell proved that on his theory electromagnetic disturbances are propagated as a wave motion through the dielectric, while Lord Kelvin in 1853 (*Phil. Mag.* [4] 5, p. 393) proved from electromagnetic theory that the discharge of a condenser is oscillatory, a result which Feddersen (*Pogg. Ann.* 103, p. 69, &c.) verified by a beautiful series of experiments. The oscillating discharge of a condenser had been inferred by Henry as long ago as 1842 from his experiments on the magnetization produced in needles by the discharge of a condenser. From these two results it follows that electric waves must be passing through the dielectric surrounding a condenser in the act of discharging, but it was not until 1887 that the existence of such waves was demonstrated by direct experiment. This great step was made by Hertz (*Wied. Ann.* 34, pp. 155, 551, 609; *Ausbreitung der elektrischen Kraft*, Leipzig, 1892), whose experiments on this subject form one of the greatest contributions ever made to experimental physics. The difficulty which had stood in the way of the observations of these waves was the absence of any method of detecting electrical and magnetic forces, reversed some millions of times per second, and only lasting for an exceedingly short time. This was removed by Hertz, who showed that such forces would produce small sparks between pieces of metal very nearly in contact, and that these sparks were sufficiently regular to be used to detect electric waves and to investigate their properties. Other and more delicate methods have subsequently been discovered, but the results obtained by Hertz with his detector were of such signal importance, that we shall begin our account of experiments on these waves by a description of some of Hertz's more fundamental experiments.

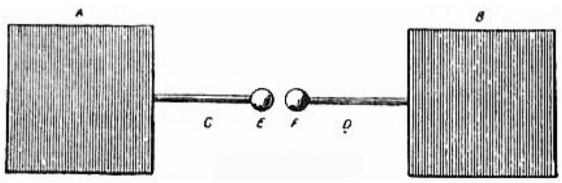

FIG. 1.

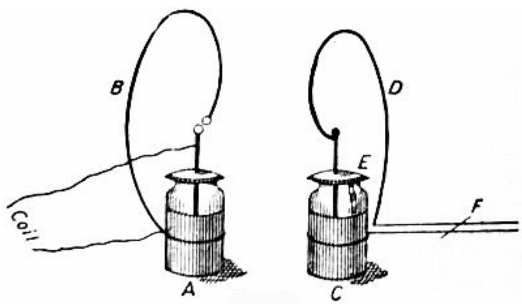

FIG. 2.

To produce the waves Hertz used two forms of vibrator. The first is represented in fig. 1. A and B are two zinc plates about 40 cm. square; to these brass rods, C, D, each about 30 cm. long, are soldered, terminating in brass balls E and F. To get good results it is necessary that these balls should be very brightly polished, and as they get roughened by the sparks which pass between them it is necessary to repolish them at short intervals; they should be shaded from light and from sparks, or other source of ultra-violet light. In order to excite the waves, C and D are connected to the two poles of an induction coil; sparks cross the air-gap which becomes a conductor, and the charges on the plates oscillate backwards and forwards like the charges on the coatings of a Leyden jar when it is short-circuited. The object of polishing the balls and screening off light is to get a sudden and sharp discharge; if the balls are rough there will be sharp points from which the charge will gradually leak, and the discharge will not be abrupt enough to start electrical vibrations, as these have an exceedingly short period. From the open form of this vibrator we should expect the radiation to be very large and the rate of decay of the amplitude very rapid. Bjerknes (*Wied. Ann.* 44, p. 74) found that the amplitude fell to 1/e of the original value, after a time 4T where T was the period of the electrical vibrations. Thus after a few vibrations the amplitude becomes inappreciable. To detect the waves produced by this vibrator Hertz used a piece of copper wire bent into a circle, the ends being furnished with two balls, or a ball and a point connected by a screw, so that the distance between them admitted of very fine adjustment. The radius of the circle for use with the vibrator just described was 35 cm., and was so chosen that the free period of the detector might be the same as that of the vibrator, and the effects in it increased by resonance. It is evident, however, that with a primary system as greatly damped as the vibrator used by Hertz, we could not expect very marked resonance effects, and as a matter of fact the accurate timing of vibrator and detector in this case is not very important. With electrical vibrators which can maintain a large number of vibrations, resonance effects are very striking, as is beautifully shown by the following experiment due to Lodge (*Nature*, 41, p. 368), whose researches have greatly advanced our knowledge of electric waves. A and C (fig. 2) are two Leyden jars, whose inner and outer coatings are connected by wires, B and D, bent so as to include a considerable area. There is an air-break in the circuit connecting the inside and outside of one of the jars, A, and electrical oscillations are started in A by joining the inside and outside with the terminals of a coil or electrical machine. The circuit in the jar C is provided with a sliding piece, F, by means of which the 204
self-induction of the discharging circuit, and, therefore, the time of an electrical oscillation of the jar, can be adjusted. The inside and outside of this jar are put almost, but not quite, into electrical contact by means of a piece of tin-foil, E, bent over the lip of the jar. The jars are placed face to face so that the circuits B and D are parallel to each other, and approximately at right angles to the line joining their centres. When the electrical machine is in action sparks pass across the air-break in the circuit in A, and by moving the slider F it is possible to find one position for it in which sparks pass from the inside to the outside of C across the tin-foil, while when the slider is moved a short distance on either side of this position the sparks cease.

Hertz found that when he held his detector in the neighbourhood of the vibrator minute sparks passed between the balls. These sparks were not stopped when a large plate of non-conducting substance, such as the wall of a room, was interposed between the vibrator and detector, but a large plate of very thin metal stopped them completely.

To illustrate the analogy between electric waves and waves of light Hertz found another form of apparatus more convenient. The vibrator consisted of two equal brass cylinders, 12 cm. long and 3 cm. in diameter, placed with their axes coincident, and in the focal line of a large zinc parabolic mirror about 2 m. high, with a focal length of 12.5 cm. The ends of the cylinders nearest each other, between which the sparks passed, were carefully polished. The detector, which was placed in the focal line of an equal parabolic mirror, consisted of two lengths of wire, each having a straight piece about 50 cm. long and a curved piece about 15 cm. long bent round at right angles so as to pass through the back of the mirror. The ends which came through the mirror were connected with a spark micrometer, the sparks being observed from behind the mirror. The mirrors are shown, in fig. 3.

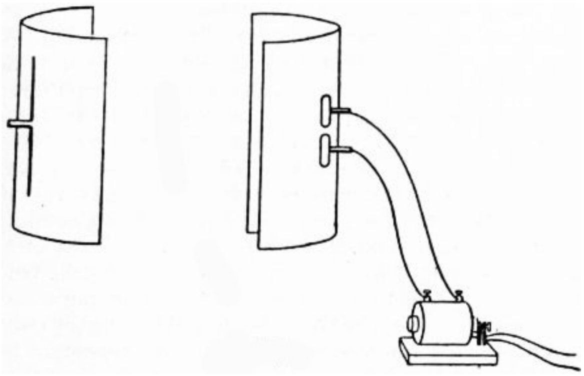

Fig. 3.

§ 2. *Reflection and Refraction.*—To show the reflection of the waves Hertz placed the mirrors side by side, so that their openings looked in the same direction, and their axes converged at a point about 3 m. from the mirrors. No sparks were then observed in the detector when the vibrator was in action. When, however, a large zinc plate about 2 m. square was placed at right angles to the line bisecting the angle between the axes of the mirrors sparks became visible, but disappeared again when the metal plate was twisted through an angle of about 15° to either side. This experiment showed that electric waves are reflected, and that, approximately at any rate, the angle of incidence is equal to the angle of reflection. To show refraction Hertz used a large prism made of hard pitch, about 1.5 m. high, with a slant side of 1.2 m. and an angle of 30°. When the waves from the vibrator passed through this the sparks in the detector were not excited when the axes of the two mirrors were parallel, but appeared when the axis of the mirror containing the detector made a certain angle with the axis of that containing the vibrator. When the system was adjusted for minimum deviation the sparks were most vigorous when the angle between the axes of the mirrors was 22°. This corresponds to an index of refraction of 1.69.

§ 3. *Analogy to a Plate of Tourmaline.*—If a screen be made by winding wire round a large rectangular framework, so that the turns of the wire are parallel to one pair of sides of the frame, and if this screen be interposed between the parabolic mirrors when placed so as to face each other, there will be no sparks in the detector when the turns of the wire are parallel to the focal lines of the mirror; but if the frame is turned through a right angle so that the wires are perpendicular to the focal lines of the mirror the sparks will recommence. If the framework is substituted for the metal plate in the experiment on the reflection of electric waves, sparks will appear in the detector when the wires are parallel to the focal lines of the mirrors, and will disappear when the wires are at right angles to these lines. Thus the framework reflects but does not transmit the waves when the electric force in them is parallel to the wires, while it transmits but does not reflect waves in which the electric force is at right angles to the wires. The wire framework behaves towards the electric waves exactly as a plate of tourmaline does to waves of light. Du Bois and Rubens (*Wied. Ann.* 49, p. 593), by using a framework wound with very fine wire placed very close together, have succeeded in polarizing waves of radiant heat, whose wave length, although longer than that of ordinary light, is very small compared with that of electric waves.

§ 4. *Angle of Polarization.*—When light polarized at right angles to the plane of incidence falls on a refracting substance at an angle $\tan^{-1}\mu$, where μ is the refractive index of the substance, all the light is refracted and none reflected; whereas when light is polarized in the plane of incidence, some of the light is always reflected whatever the angle of incidence. Trouton (*Nature*, 39, p. 391) showed that similar effects take place with electric waves. From a paraffin wall 3 ft. thick, reflection always took place when the electric force in the incident wave was at right angles to the plane of incidence, whereas at a certain angle of incidence there was no reflection when the vibrator was turned, so that the electric force was in the plane of incidence. This shows that on the electromagnetic theory of light the electric force is at right angles to the plane of polarization.

§ 5. *Stationary Electrical Vibrations.*—Hertz (*Wied. Ann.* 34, p. 609) made his experiments on these in a large room about 15 m. long. The vibrator, which was of the type first described, was placed at one end of the room, its plates being parallel to the wall, at the other end a piece of sheet zinc about 4 m. by 2 m. was placed vertically against the wall. The detector—the circular ring previously described—was held so that its plane was parallel to the metal plates of the vibrator, its centre on the line at right angles to the metal plate bisecting at right angles the spark gap of the vibrator, and with the spark gap of the detector

parallel to that of the vibrator. The following effects were observed when the detector was moved about. When it was close up to the zinc plate there were no sparks, but they began to pass feebly as soon as it was moved forward a little way from the plate, and increased rapidly in brightness until it was about 1.8 m. from the plate, when they attained their maximum. When its distance was still further increased they diminished in brightness, and vanished again at a distance of about 4 m. from the plate. When the distance was still further increased they reappeared, attained another maximum, and so on. They thus exhibited a remarkable periodicity similar to that which occurs when stationary vibrations are produced by the interference of direct waves with those reflected from a surface placed at right angles to the direction of propagation. Similar periodic alterations in the spark were observed by Hertz when the waves, instead of passing freely through the air and being reflected by a metal plate at the end of the room, were led along wires, as in the arrangement shown in fig. 4. L and K are metal plates placed parallel to the plates of the vibrator, long parallel wires being attached to act as guides to the waves which were reflected from the isolated end. (Hertz used only one plate and one wire, but the double set of plates and wires introduced 205
by Sarasin and De la Rive make the results more definite.) In this case the detector is best placed so that its plane is at right angles to the wires, while the air space is parallel to the plane containing the wires. The sparks instead of vanishing when the detector is at the far end of the wire are a maximum in this position, but wax and wane periodically as the detector is moved along the wires. The most obvious interpretation of these experiments was the one given by Hertz—that there was interference between the direct waves given out by the vibrator and those reflected either from the plate or from the ends of the wire, this interference giving rise to stationary waves. The places where the electric force was a maximum were the places where the sparks were brightest, and the places where the electric force was zero were the places where the sparks vanished. On this explanation the distance between two consecutive places where the sparks vanished would be half the wave length of the waves given out by the vibrator.

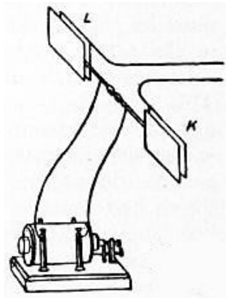

FIG. 4.

Some very interesting experiments made by Sarasin and De la Rive (*Comptes rendus*, 115, p. 489) showed that this explanation could not be the true one, since by using detectors of different sizes they found that the distance between two consecutive places where the sparks vanished depended mainly upon the size of the detector, and very little upon that of the vibrator. With small detectors they found the distance small, with large detectors, large; in fact it is directly proportional to the diameter of the detector. We can see that this result is a consequence of the large damping of the oscillations of the vibrator and the very small damping of those of the detector. Bjerknes showed that the time taken for the amplitude of the vibrations of the vibrator to sink to $1/e$ of their original value was only $4T$, while for the detector it was $500T'$, when T and T' are respectively the times of vibration of the vibrator and the detector. The rapid decay of the oscillations of the vibrator will stifle the interference between the direct and the reflected wave, as the amplitude of the direct wave will, since it is emitted later, be much smaller than that of the reflected one, and not able to annul its effects completely; while the well-maintained vibrations of the detector will interfere and produce the effects observed by Sarasin and De la Rive. To see this let us consider the extreme case in which the oscillations of the vibrator are absolutely dead-beat. Here an impulse, starting from the vibrator on its way to the reflector, strikes against the detector and sets it in vibration; it then travels up to the plate and is reflected, the electric force in the impulse being reversed by reflection. After reflection the impulse again strikes the detector, which is still vibrating from the effects of the first impact; if the phase of this vibration is such that the reflected impulse tends to produce a current round the detector in the same direction as that which is circulating from the effects of the first impact, the sparks will be increased, but if the reflected impulse tends to produce a current in the opposite direction the sparks will be diminished. Since the electric force is reversed by reflection, the greatest increase in the sparks will take place when the impulse finds, on its return, the detector in the opposite phase to that in which it left it; that is, if the time which has elapsed between the departure and return of the impulse is equal to an odd multiple of half the time of vibration of the detector. If d is the distance of the detector from the reflector when the sparks are brightest, and V the velocity of propagation of electromagnetic disturbance, then $2d/V = (2n + 1)(T'/2)$; where n is an integer and T' the time of vibration of the detector, the distance between two spark maxima will be $VT'/2$, and the places where the sparks are a minimum will be midway between the maxima. Sarasin and De la Rive found that when the same detector was used the distance between two spark maxima was the same with the waves through air reflected from a metal plate and with those guided by wires and reflected from the free ends of the wire, the inference being that the velocity of waves along wires is the same as that through the air. This result, which follows from Maxwell’s theory, when the wires are not too fine, had been questioned by Hertz on account of some of his experiments on wires.

§ 6. *Detectors.*—The use of a detector with a period of vibration of its own thus tends to make the experiments more complicated, and many other forms of detector have been employed by subsequent experimenters. For example, in place of the sparks in air the luminous discharge through a rarefied gas has been used by Dragoumis, Lecher (who used tubes without electrodes laid across the wires in an arrangement resembling that shown in fig. 7) and Arons. A tube containing neon at a low pressure is especially suitable for this purpose. Zehnder (*Wied. Ann.* 47, p. 777) used an exhausted tube to which an external electromotive force almost but not quite sufficient of itself to produce a discharge was applied; here the additional electromotive force due to the waves was sufficient to start the discharge. Detectors depending on the heat produced by the rapidly alternating currents have been used by Paalzow and Rubens, Rubens and Ritter, and I. Klemenčič. Rubens measured the heat produced by a bolometer arrangement, and Klemenčič used a thermo-electric method for the same purpose; in consequence of the great increase in the sensitiveness of galvanometers these methods are now very frequently resorted to. Boltzmann used an electroscope as a detector. The spark gap consisted of a ball and a point, the ball being connected with the electroscope and the point with a battery of 200 dry cells. When the spark passed the cells charged up the electroscope. Ritter utilized the contraction of a frog’s leg as a detector, Lucas and Garrett the explosion produced by the sparks in an explosive mixture of hydrogen and oxygen; while Bjerknes and Franke used the mechanical attraction between oppositely charged conductors. If the two sides of the spark gap are connected with the two pairs of

quadrants of a very delicate electrometer, the needle of which is connected with one pair of quadrants, there will be a deflection of the electrometer when the detector is struck by electric waves. A very efficient detector is that invented by E. Rutherford (*Trans. Roy. Soc.* A. 1897, 189, p. 1); it consists of a bundle of fine iron wires magnetized to saturation and placed inside a small magnetizing coil, through which the electric waves cause rapidly alternating currents to pass which demagnetize the soft iron. If the instrument is used to detect waves in air, long straight wires are attached to the ends of the demagnetizing coil to collect the energy from the field; to investigate waves in wires it is sufficient to make a loop or two in the wire and place the magnetized piece of iron inside it. The amount of demagnetization which can be observed by the change in the deflection of a magnetometer placed near the iron, measures the intensity of the electric waves, and very accurate determinations can be made with ease with this apparatus. It is also very delicate, though in this respect it does not equal the detector to be next described, the coherer; Rutherford got indications in 1895 when the vibrator was ¾ of a mile away from the detector, and where the waves had to traverse a thickly populated part of Cambridge. It can also be used to measure the coefficient of damping of the electric waves, for since the wire is initially magnetized to saturation, if the direction of the current when it first begins to flow in the magnetizing coil is such as to tend to increase the magnetization of the wire, it will produce no effect, and it will not be until the current is reversed that the wire will lose some of its magnetization. The effect then gives the measure of the intensity half a period after the commencement of the waves. If the wire is put in the coil the opposite way, *i.e.* so that the magnetic force due to the current begins at once to demagnetize the wire, the demagnetization gives a measure of the initial intensity of the waves. Comparing this result with that obtained when the wires were reversed, we get the coefficient of damping. A very convenient detector of electric waves is the one discovered almost simultaneously by Fessenden (*Electrotech. Zeits.*, 1903, 24, p. 586) and Schlömilch (*ibid.* p. 959). This consists of an electrolytic cell in which one of the electrodes is an exceedingly fine point. The electromotive force in the circuit is small, and there is large polarization in the circuit with only a small current. When the circuit is struck by electric 206
waves there is an increase in the currents due to the depolarization of the circuit. If a galvanometer is in the circuit, the increased deflection of the instrument will indicate the presence of the waves.

§ 7. *Coherers.*—The most sensitive detector of electric waves is the “coherer,” although for metrical work it is not so suitable as that just described. It depends upon the fact discovered by Branly (*Comptes rendus*, 111, p. 785; 112, p. 90) that the resistance between loose metallic contacts, such as a pile of iron turnings, diminishes when they are struck by an electric wave. One of the forms made by Lodge (*The Work of Hertz and some of his Successors*, 1894) on this principle consists simply of a glass tube containing iron turnings, in contact with which are wires led into opposite ends of the tube. The arrangement is placed in series with a galvanometer (one of the simplest kind will do) and a battery; when the iron turnings are struck by electric waves their resistance is diminished and the deflection of the galvanometer is increased. Thus the deflection of the galvanometer can be used to indicate the arrival of electric waves. The tube must be tapped between each experiment, and the deflection of the galvanometer brought back to about its original value. This detector is marvellously delicate, but not metrical, the change produced in the resistance depending upon so many things besides the intensity of the waves that the magnitude of the galvanometer deflection is to some extent a matter of chance. Instead of the iron turnings we may use two iron wires, one resting on the other; the resistance of this contact will be altered by the incidence of the waves. To get greater regularity Bose uses, instead of the iron turnings, spiral springs, which are pushed against each other by means of a screw until the most sensitive state is attained. The sensitiveness of the coherer depends on the electromotive force put in the galvanometer circuit. Very sensitive ones can be made by using springs of very fine silver wire coated electrolytically with nickel. Though the impact of electric waves generally produces a diminution of resistance with these loose contacts, yet there are exceptions to the rule. Thus Branly showed that with lead peroxide, PbO_2, there is an increase in resistance. Aschkinass proved the same to be true with copper sulphide, CuS; and Bose showed that with potassium there is an increase of resistance and great power of self-recovery of the original resistance after the waves have ceased. Several theories of this action have been proposed. Branly (*Lumière électrique*, 40, p. 511) thought that the small sparks which certainly pass between adjacent portions of metal clear away layers of oxide or some other kind of non-conducting film, and in this way improve the contact. It would seem that if this theory is true the films must be of a much more refined kind than layers of oxide or dirt, for the coherer effect has been observed with clean non-oxidizable metals. Lodge explains the effect by supposing that the heat produced by the sparks fuses adjacent portions of metal into contact and hence diminishes the resistance; it is from this view of the action that the name coherer is applied to the detector. Auerbeck thought that the effect was a mechanical one due to the electrostatic attractions between the various small pieces of metal. It is probable that some or all of these causes are at work in some cases, but the effects of potassium make us hesitate to accept any of them as the complete explanation. Blanc (*Ann. chim. phys.*, 1905, [8] 6, p. 5), as the result of a long series of experiments, came to the conclusion that coherence is due to pressure. He regarded the outer layers as different from the mass of the metal and having a much greater specific resistance. He supposed that when two pieces of metal are pressed together the molecules diffuse across the surface, modifying the surface layers and increasing their conductivity.

§ 8. *Generators of Electric Waves.*—Bose (*Phil. Mag.* 43, p. 55) designed an instrument which generates electric waves with a length of not more than a centimetre or so, and therefore allows their properties to be demonstrated with apparatus of moderate dimensions. The waves are excited by sparking between two platinum beads carried by jointed electrodes; a platinum sphere is placed between the beads, and the distance between the beads and the sphere can be adjusted by bending the electrodes. The diameter of the sphere is 8 mm., and the wave length of the shortest electrical waves generated is said to be about 6 mm. The beads are connected with the terminals of a small induction coil, which, with the battery to work it and the sparking arrangement, are enclosed in a metal box, the radiation passing out through a metal tube opposite to the spark gap. The ordinary vibrating break of the coil is not used, a single spark made by making and breaking the circuit by means of a button outside the box being employed instead. The detector is one of the spiral spring coherers previously described; it is shielded from external disturbance by being enclosed in a metal box provided with a funnel-shaped opening to admit the radiation. The wires leading from the coherers to the galvanometer are also surrounded by metal tubes to protect them from stray radiation. The radiating apparatus and the receiver are mounted on stands sliding in an optical bench. If a parallel beam of radiation is required, a cylindrical lens of ebonite or sulphur is mounted in a tube fitting on to the radiator tube and stopped by a guide when the spark is at the principal focal line of the lens. For experiments requiring angular measurements a spectrometer circle is mounted on one of the sliding stands, the receiver being carried on a radial arm and pointing to the centre of the circle. The arrangement is represented in fig. 5.

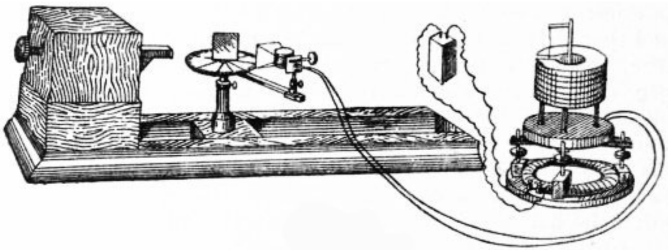

Fig. 5.

With this apparatus the laws of reflection, refraction and polarization can readily be verified, and also the double refraction of crystals, and of bodies possessing a fibrous or laminated structure such as jute or books. (The double refraction of electric waves seems first to have been observed by Righi, and other researches on this subject have been made by Garbasso and Mack.) Bose showed the rotation of the plane of polarization by means of pieces of twisted jute rope; if the pieces were arranged so that their twists were all in one direction and placed in the path of the radiation, they rotated the plane of polarization in a direction depending upon the direction of twist; if they were mixed so that there were as many twisted in one direction as the other, there was no rotation.

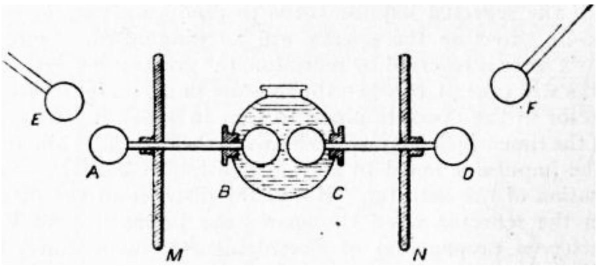

Fig. 6.

A series of experiments showing the complete analogy between electric and light waves is described by Righi in his book *L'Ottica delle oscillazioni elettriche*. Righi's exciter, which is especially convenient when large statical electric machines are used instead of induction coils, is shown in fig. 6. E and F are balls connected with the terminals of the machine, and AB and CD are conductors insulated from each other, the ends B, C, between which the sparks pass, being immersed in vaseline oil. The period of the vibrations given out by the system is adjusted by means of metal plates M and N attached to AB and CD. When the waves are produced by induction coils or by electrical machines the intervals between the emission of different sets of waves occupy by far the largest part of the time. Simon (*Wied. Ann.*, 1898, 64, p. 293; *Phys. Zeit.*, 1901, 2, p. 253), Duddell (*Electrician*, 1900, 46, p. 269) and Poulsen (*Electrotech. Zeits.*, 1906, 27, p. 1070) reduced these intervals very considerably by using the electric arc to excite the waves, and in this way produced electrical waves possessing great energy. In these methods the terminals between which the arc is passing are connected through coils with self-induction L to the plates of a condenser of capacity C. The arc is not steady, but is continually varying. This is especially the case when it passes through hydrogen. These variations excite vibrations with a period 2π√(LC) in the circuit containing the capacity of the self-induction. By this method Duddell produced waves with a frequency of 40,000. Poulsen, who cooled the terminals of the arc, produced waves with a frequency of 1,000,000, while Stechodro (*Ann. der Phys.* 27, p. 225) claims to have produced waves with three hundred times this frequency, *i.e.* having a wave length of about a metre. When the self-induction and capacity are large so that the frequency comes within the limits of the frequency 207
of audible notes, the system gives out a musical note, and the arrangement is often referred to as the singing arc.

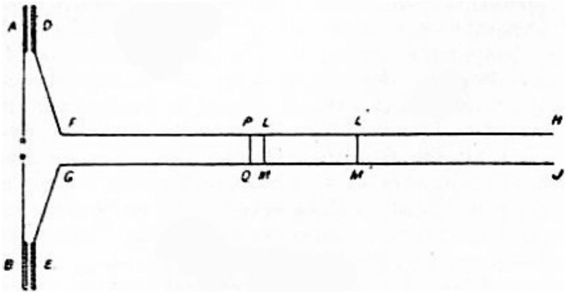

Fig. 7.

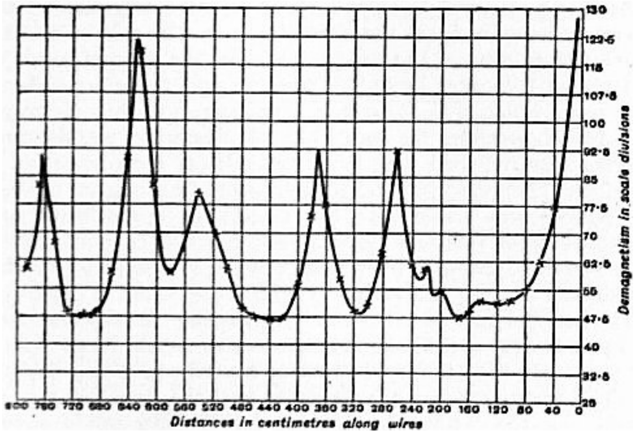

FIG. 8.

§ *9. Waves in Wires.*—Many problems on electric waves along wires can readily be investigated by a method due to Lecher (*Wied. Ann.* 41, p. 850), and known as Lecher’s bridge, which furnishes us with a means of dealing with waves of a definite and determinable wave-length. In this arrangement (fig. 7) two large plates A and B are, as in Hertz’s exciter, connected with the terminals of an induction coil; opposite these and insulated from them are two smaller plates D, E, to which long parallel wires DFH, EGJ are attached. These wires are bridged across by a wire LM, and their farther ends H, J, may be insulated, or connected together, or with the plates of a condenser. To detect the waves in the circuit beyond the bridge, Lecher used an exhausted tube placed across the wires, and Rubens a bolometer, but Rutherford’s detector is the most convenient and accurate. If this detector is placed in a fixed position at the end of the circuit, it is found that the deflections of this detector depend greatly upon the position of the bridge LM, rising rapidly to a maximum for some positions, and falling rapidly away when the bridge is displaced. As the bridge is moved from the coil end towards the detector the deflections show periodic variations, such as are represented in fig. 8 when the ordinates represent the deflections of the detector and the abscissae the distance of the bridge from the ends D, E. The maximum deflections of the detector correspond to the positions in which the two circuits DFLMGE, HLMJ (in which the vibrations are but slightly damped) are in resonance. For since the self-induction and resistance of the bridge LM is very small compared with that of the circuit beyond, it follows from the theory of circuits in parallel that only a small part of the current will in general flow round the longer circuit; it is only when the two circuits DFLMGE, HLMJ are in resonance that a considerable current will flow round the latter. Hence when we get a maximum effect in the detector we know that the waves we are dealing with are those corresponding to the free periods of the system HLMJ, so that if we know the free periods of this circuit we know the wave length of the electric waves under consideration. Thus if the ends of the wires H, J are free and have no capacity, the current along them must vanish at H and J, which must be in opposite electric condition. Hence half the wave length must be an odd submultiple of the length of the circuit HLMJ. If H and J are connected together the wave length must be a submultiple of the length of this circuit. When the capacity at the ends is appreciable the wave length of the circuit is determined by a somewhat complex expression. To facilitate the determination of the wave length in such cases, Lecher introduced a second bridge L′M′, and moved this about until the deflection of the detector was a maximum; when this occurs the wave length is one of those corresponding to the closed circuit LMM′L′, and must therefore be a submultiple of the length of the circuit. Lecher showed that if instead of using a single wire LM to form the bridge, he used two parallel wires PQ, LM, placed close together, the currents in the further circuit were hardly appreciably diminished when the main wires were cut between PL and QM. Blondlot used a modification of this apparatus better suited for the production of short waves. In his form (fig. 9) the exciter consists of two semicircular arms connected with the terminals of an induction coil, and the long wires, instead of being connected with the small plates, form a circuit round the exciter.

As an example of the use of Lecher’s arrangement, we may quote Drude’s application of the method to find the specific induction capacity of dielectrics under electric oscillations of varying frequency. In this application the ends of the wire are connected to the plates of a condenser, the space between whose plates can be filled with the liquid whose specific inductive capacity is required, and the bridge is moved until the detector at the end of the circuit gives the maximum deflection. Then if λ is the wave length of the waves, λ is the wave length of one of the free vibrations of the system HLMJ; hence if C is the capacity of the condenser at the end in electrostatic measure we have

$$\frac{\cot\frac{2\pi l}{\lambda}}{\frac{2\pi l}{\lambda}} = \frac{C}{C'l}$$

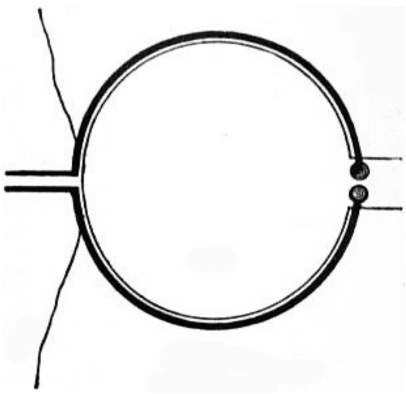

FIG. 9.

where l is the distance of the condenser from the bridge and C′ is the capacity of unit length of the wire. In the condenser part of the lines of force will pass through air and part through the dielectric; hence C will be of the form $C_0 + KC_1$ where K is the specific inductive capacity of the dielectric. Hence if l is the distance of maximum deflection when the dielectric is replaced by air, *l′* when filled with a dielectric whose specific inductive capacity is known to be K′, and l″ the distance when filled with the dielectric whose specific inductive capacity is required, we easily see that—

$$\frac{\cot\frac{2\pi l}{\lambda} - \cot\frac{2\pi l'}{\lambda}}{2\pi l \qquad\qquad 2\pi l''} = \frac{1 - K'}{1 - K}$$

$$\cot \frac{}{\lambda} - \cot \frac{}{\lambda}$$

an equation by means of which K can be determined. It was in this way that Drude investigated the specific inductive capacity with varying frequency, and found a falling off in the specific inductive capacity with increase of frequency when the dielectrics contained the radicle OH. In another method used by him the wires were led through long tanks filled with the liquid whose specific inductive capacity was required; the velocity of propagation of the electric waves along the wires in the tank being the same as the velocity of propagation of an electromagnetic disturbance through the liquid filling the tank, if we find the wave length of the waves along the wires in the tank, due to a vibration of a given frequency, and compare this with the wave lengths corresponding to the same frequency when the wires are surrounded by air, we obtain the velocity of propagation of electromagnetic disturbance through the fluid, and hence the specific inductive capacity of the fluid.

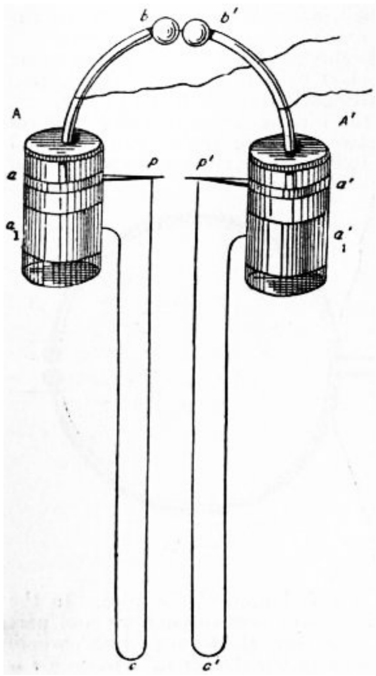

FIG. 10.

§ 10. *Velocity of Propagation of Electromagnetic Effects through Air.*—The experiments of Sarasin and De la Rive already described (see § 5) have shown that, as theory requires, the velocity of propagation of electric effects through air is the same as along wires. The same result had been arrived at by J.J. Thomson, although from the method he used greater differences between the velocities might have escaped detection than was possible by Sarasin and De la Rive’s method. The velocity of waves along wires has been directly determined by Blondlot by two different methods. In the first the detector consisted of two parallel plates about 6 cm. in diameter placed a fraction of a millimetre apart, and forming a condenser whose capacity C was determined in electromagnetic measure by Maxwell’s method. The plates were connected by a rectangular circuit whose self-induction L was calculated from the dimensions of the rectangle and the size of the wire. The time of vibration T is equal to $2\pi\sqrt{(LC)}$. (The wave length corresponding to this time is long compared with the length of the circuit, so that the use of this formula is legitimate.) This detector is placed between two parallel wires, and the waves produced by the exciter are reflected from a movable bridge. When this bridge is placed just beyond the detector vigorous sparks are observed, but as the bridge is pushed away a place is reached where the sparks disappear; this place is distance $2/\lambda$ from the detector, when λ is the wave length of the vibration given out by the detector. The sparks again disappear when the distance of the bridge from the detector is $3\lambda/4$. Thus by measuring the distance between two consecutive positions of the bridge at which the sparks disappear λ can be determined, and v, the velocity of 208
propagation, is equal to λ/T. As the means of a number of experiments Blondlot found v to be 3.02×10^{10} cm./sec., which, within the errors of experiment, is equal to 3×10^{10} cm./sec., the velocity of light. A second method used by Blondlot, and one which does not involve the calculation of the period, is as follows:—A and A′ (fig. 10) are two equal Leyden jars coated inside and outside with tin-foil. The outer coatings form two separate rings a, a_1; a′, a'_1, and the inner coatings are connected with the poles of the induction coil by means of the metal pieces b, b′. The sharply pointed conductors p and p′, the points of which are about ½ mm. apart, are connected with the rings of the tin-foil a and a′, and two long copper wires pca_1, $p'c'a'_1$, 1029 cm. long, connect these points with the other rings a_1, a_1'. The rings aa′, a_1a_1', are connected by wet strings so as to charge up the jars. When a spark passes between b and b′, a spark at once passes between pp′, and this is followed by another spark when the waves travelling by the paths a_1cp, $a'_1c'p'$ reach p and p′. The time between the passage of these sparks, which is the time taken by the waves to travel 1029 cm., was observed by means of a rotating mirror, and the velocity measured in 15 experiments varied between 2.92×10^{10} and 3.03×10^{10} cm./sec., thus agreeing well with that deduced by the preceding method. Other determinations of the velocity of electromagnetic propagation have been made by Lodge and Glazebrook, and by Saunders.

On Maxwell’s electromagnetic theory the velocity of propagation of electromagnetic disturbances should equal the velocity of light, and also the ratio of the electromagnetic unit of electricity to the electrostatic unit. A large number of determinations of this ratio have been made:—

Observer.	Date.	Ratio 10^{10} ×.
Klemenčič	1884	3.019 cm./sec.
Himstedt	1888	3.009 cm./sec.
Rowland	1889	2.9815 cm./sec.
Rosa	1889	2.9993 cm./sec.
J.J. Thomson and Searle	1890	2.9955 cm./sec.
Webster	1891	2.987 cm./sec.
Pellat	1891	3.009 cm./sec.
Abraham	1892	2.992 cm./sec.
Hurmuzescu	1895	3.002 cm./sec.
Rosa	1908	2.9963 cm./sec.

The mean of these determinations is 3.001×10^{10} cm./sec., while the mean of the last five determinations of the velocity of light in air is given by Himstedt as 3.002×10^{10} cm./sec. From these experiments we conclude that the velocity of propagation of an electromagnetic disturbance is equal to the velocity of light, and to the velocity required by Maxwell’s theory.

In experimenting with electromagnetic waves it is in general more difficult to measure the period of the oscillations than their wave length. Rutherford used a method by which the period of the vibration can easily be determined; it is based upon the theory of the distribution of alternating currents in two circuits ACB, ADB in parallel. If A and B are respectively the maximum currents in the circuits ACB, ADB, then

$$\frac{A}{B} = \sqrt{\frac{S^2 + (N - M)^2p^2}{R^2 + (L - M)^2p^2}}$$

when R and S are the resistances, L and N the coefficients of self-induction of the circuits ACB, ADB respectively, M the coefficient of mutual induction between the circuits, and p the frequency of the currents. Rutherford detectors were placed in the two circuits, and the circuits adjusted until they showed that A = B; when this is the case

$$p^2 = \frac{R^2 - S^2}{N^2 - L^2 - 2M(N - L)}.$$

If we make one of the circuits, ADB, consist of a short length of a high liquid resistance, so that S is large and N small, and the other circuit ACB of a low metallic resistance bent to have considerable self-induction, the preceding equation becomes approximately p = S/L, so that when S and L are known p is readily determined.

(J. J. T.)

ELECTROCHEMISTRY. The present article deals with processes that involve the electrolysis of aqueous solutions, whilst those in which electricity is used in the manufacture of chemical products at furnace temperatures are treated under Electrometallurgy, although, strictly speaking, in some cases (*e.g.* calcium carbide and phosphorus manufacture) they are not truly metallurgical in character. For the theory and elemental laws of electro-deposition see Electrolysis; and for the construction and use of electric generators see Dynamo and Battery: *Electric*. The importance of the subject may be gauged by the fact that all the aluminium, magnesium, sodium, potassium, calcium carbide, carborundum and artificial graphite, now placed on the market, is made by electrical processes, and that the use of such processes for the refining of copper and silver, and in the manufacture of phosphorus, potassium chlorate and bleach, already pressing very heavily on the older non-electrical systems, is every year extending. The convenience also with which the energy of waterfalls can be converted into electric energy has led to the introduction of chemical industries into countries and districts where, owing to the absence of coal, they were previously unknown. Norway and Switzerland have become important producers of chemicals, and pastoral districts such as those in which Niagara or Foyers are situated manufacturing centres. In this way the development of the electrochemical industry is in a marked degree altering the distribution of trade throughout the world.

Electrolytic Refining of Metals.—The principle usually followed in the electrolytic refining of metals is to cast the impure metal into plates, which are exposed as anodes in a suitable solvent, commonly a salt of the metal under treatment. On passing a current of electricity, of which the volume and pressure are adjusted to the conditions of the electrolyte and electrodes, the anode slowly dissolves, leaving the insoluble impurities in the form of a sponge, if the proportion be considerable, but otherwise as a mud or slime which becomes detached from the anode surface and must be prevented from coming into contact with the cathode. The metal to be refined passing into solution is concurrently deposited at the cathode. Soluble impurities which are more electro-negative than the metal under treatment must, if present, be removed by a preliminary process, and the voltage and other conditions must be so selected that none of the more electro-positive metals are co-deposited with the metal to be refined. From these and other considerations it is obvious that (1) the electrolyte must be such as will freely dissolve the metal to be refined; (2) the electrolyte must be able to dissolve the major portion of the anode, otherwise the mass of insoluble matter on the outer layer will prevent access of electrolyte to the core, which will thus escape refining; (3) the electrolyte should, if possible, be incapable of dissolving metals more electro-negative than that to be refined; (4) the proportion of soluble electro-positive impurities must not be excessive, or these substances will accumulate too rapidly in the solution and necessitate its frequent purification; (5) the current density must be so adjusted to the strength of the solution and to other conditions that no relatively electro-positive metal is deposited, and that the cathode deposit is physically suitable for subsequent treatment; (6) the current density should be as high as is consistent with the production of a pure and sound deposit, without undue expense of voltage, so that the operation may be rapid and the "turnover" large; (7) the electrolyte should be as good a conductor of electricity as possible, and should not, ordinarily, be altered chemically by exposure to air; and (8) the use of porous partitions should be avoided, as they increase the resistance and usually require frequent renewal. For details of the practical methods see Gold; Silver; Copper and headings for other metals.

Electrolytic Manufacture of Chemical Products.—When an aqueous solution of the salt of an alkali metal is electrolysed, the metal reacts with the water, as is well known, forming caustic alkali, which dissolves in the solution, and hydrogen, which comes off as a gas. So early as 1851 a patent was taken out by Cooke for the production of caustic alkali without the use of a separate current, by immersing iron and copper plates on opposite sides of a porous (biscuit-ware) partition in a suitable cell, containing a solution of the salt to be electrolysed, at 21°-65° C. (70°-150° F.). The solution of the iron anode was intended to afford the necessary energy. In the same year another patent was granted to C. Watt for a similar process, involving the employment of an externally generated current. When an alkaline chloride, say sodium chloride, is electrolysed with one electrode immersed in a porous cell, while caustic soda is formed at the cathode, chlorine is deposited at the anode. If the latter be insoluble, the gas diffuses into the solution and, when this becomes saturated, escapes into the air. If, however, no porous division be used to prevent the intermingling by diffusion of the anode and cathode solutions, a complicated set of subsidiary reactions takes place. The chlorine reacts with the caustic soda, forming sodium hypochlorite, and this in turn, with an excess of chlorine and at higher temperatures, becomes for the most part converted into chlorate, whilst any simultaneous electrolysis of a hydroxide or water and a chloride (so that hydroxyl and chlorine are simultaneously liberated at the anode) also produces oxygen-chlorine compounds direct. At the same time, 209 the diffusion of these compounds into contact with the cathode leads to a partial reduction to chloride, by the removal of combined oxygen by the instrumentality of the hydrogen there evolved. In proportion as the original chloride is thus reproduced, the efficiency of the process is of course diminished. It is obvious that, with suitable methods and apparatus, the electrolysis of alkaline chlorides may be made to yield chlorine, hypochlorites (bleaching liquors), chlorates or caustic alkali, but that great care must be exercised if any of these products is to be obtained pure and with economy. Many patents have been taken out in this branch of electrochemistry, but it is to be remarked that that granted to C. Watt traversed the whole of the ground. In his process a current was passed through a tank divided into two or three cells by porous partitions, hoods and tubes were arranged to carry off chlorine and hydrogen respectively, and the whole was heated to 120° F. by a steam jacket when caustic alkali was being made. Hypochlorites were made, at ordinary temperatures, and chlorates at higher temperatures, in a cell without a partition in which the cathode was placed horizontally immediately above the anode, to favour the mixing of the ascending chlorine with the descending caustic solution.

The relation between the composition of the electrolyte and the various conditions of current-density, temperature and the like has been studied by F. Oettel (*Zeitschrift f. Elektrochem.*, 1894, vol. i. pp. 354 and 474) in connexion with the production of hypochlorites and chlorates in tanks without diaphragms, by C. Häussermann and W. Naschold (*Chemiker Zeitung*, 1894, vol. xviii. p. 857) for their production in cells with porous diaphragms, and by F. Haber and S. Grinberg (*Zeitschrift f. anorgan. Chem.*, 1898, vol. xvi. pp. 198, 329, 438) in connexion with the electrolysis of hydrochloric acid. Oettel, using a 20% solution of potassium chloride, obtained the best yield of hypochlorite with a high current-density, but as soon as 1¼% of bleaching chlorine (as hypochlorite) was present, the formation of chlorate commenced. The yield was at best very low as compared with that theoretically possible. The best yield of chlorate was obtained when from 1 to 4% of caustic potash was present. With high current-density, heating the solution tended to increase the proportion of chlorate to hypochlorite, but as the proportion of water decomposed is then higher, the amount of chlorine produced must be less and the total chlorine efficiency lower. He also traced a connexion between alkalinity, temperature and current-density, and showed that these conditions should be mutually adjusted. With a current-density of 130 to 140 amperes per sq. ft., at 3 volts, passing between platinum electrodes, he attained to a current-efficiency of 52%, and each (British) electrical horse-power hour was equivalent to a production of 1378.5 grains of potassium chlorate. In other words, each pound of chlorate would require an expenditure of nearly 5.1 e.h.p. hours. One of the earliest of the more modern processes was that of E. Hermite, which consisted in the production of bleach-liquors by the electrolysis (according to the 1st edition of the 1884 patent) of magnesium or calcium chloride between platinum anodes carried in wooden frames, and zinc cathodes. The solution, containing hypochlorites and chlorates, was then applied to the bleaching of linen, paper-pulp or the like, the solution being used over and over again. Many modifications have been patented by Hermite, that of 1895 specifying the use of platinum gauze anodes, held in ebonite or other frames. Rotating zinc cathodes were used, with scrapers to prevent the accumulation of a layer of insoluble magnesium compounds, which would otherwise increase the electrical resistance beyond reasonable limits. The same inventor has patented the application of electrolysed chlorides to the purification of starch by the oxidation of less stable organic bodies, to the bleaching of oils, and to the purification of coal gas, spirit and other substances. His system for the disinfection of sewage and similar matter by the electrolysis of chlorides, or of sea-water, has been tried, but for the most part abandoned on the score of expense. Reference may be made to papers written in the early days of the process by C.F. Cross and E.J. Bevan (*Journ. Soc. Chem. Industry*, 1887, vol. vi. p. 170, and 1888, vol. vii. p. 292), and to later papers by P. Schoop (*Zeitschrift f. Elektrochem.*, 1895, vol. ii. pp. 68, 88, 107, 209, 289).

E. Kellner, who in 1886 patented the use of cathode (caustic soda) and anode (chlorine) liquors in the manufacture of cellulose from wood-fibre, and has since evolved many similar processes, has produced an apparatus that has been largely used. It consists of a stoneware tank with a thin sheet of platinum-iridium alloy at either end forming the primary electrodes, and between them a number of glass plates reaching nearly to the bottom, each having a platinum gauze sheet on either side; the two sheets belonging to each plate are in metallic connexion, but insulated from all the others, and form intermediary or bi-polar electrodes. A 10-12% solution of sodium chloride is caused to flow upwards through the apparatus and to overflow into troughs, by which it is conveyed (if necessary through a cooling apparatus) back to the circulating pump. Such a plant has been reported as giving 0.229 gallon of a liquor containing 1% of available chlorine per kilowatt hour, or 0.171 gallon per e.h.p. hour. Kellner has also patented a “bleaching-block,” as he terms it, consisting of a frame carrying parallel plates similar in principle to those last described. The block is immersed in the solution to be bleached, and may be lifted in or out as required. O. Knöfler and Gebauer have also a system of bi-polar electrodes, mounted in a frame in appearance resembling a filter-press.

Other Electrochemical Processes.—It is obvious that electrolytic iodine and bromine, and oxygen compounds of these elements, may be produced by methods similar to those applied to chlorides (see Alkali Manufacture and Chlorates), and Kellner and others have patented processes with this end in view. *Hydrogen* and *oxygen* may also be produced electrolytically as gases, and their respective reducing and oxidizing powers at the moment of deposition on the electrode are frequently used in the laboratory, and to some extent industrially, chiefly in the field of organic chemistry. Similarly, the formation of organic halogen products may be effected by electrolytic chlorine, as, for example, in the production of *chloral* by the gradual introduction of alcohol into an anode cell in which the electrolyte is a strong solution of potassium chloride. Again, anode reactions, such as are observed in the electrolysis of the fatty acids, may be utilized, as, for example, when the radical CH_3CO_2—deposited at the anode in the electrolysis of acetic acid—is dissociated, two of the groups react to give one molecule of *ethane*, C_2H_6, and two of carbon dioxide. This, which has long been recognized as a class-reaction, is obviously capable of endless variation. Many electrolytic methods have been proposed for the purification of *sugar*; in some of them soluble anodes are used for a few minutes in weak alkaline solutions, so that the caustic alkali from the cathode reaction may precipitate chemically the hydroxide of the anode metal dissolved in the liquid, the precipitate carrying with it mechanically some of the impurities present, and thus clarifying the solution. In others the current is applied for a longer time to the original sugar-solution with insoluble (*e.g.* carbon) anodes. F. Peters has found that with these methods the best results are obtained when ozone is employed in addition to electrolytic oxygen. Use has been made of electrolysis in *tanning* operations, the current being passed through the tan-liquors containing the hides. The current, by endosmosis, favours the passage of the solution into the hide-substance, and at the same time appears to assist the chemical combinations there occurring; hence a great reduction in the time required for the completion of the process. Many patents have been taken out in this direction, one of the best known being that of Groth, experimented upon by S. Rideal and A.P. Trotter (*Journ. Soc. Chem. Indust.*, 1891, vol. x. p. 425), who employed copper anodes, 4 sq. ft. in area, with current-densities of 0.375 to 1 (ranging in some cases to 7.5) ampere per sq. ft., the best results being obtained with the smaller current-densities. Electrochemical processes are often indirectly used, as for example in the Villon process (*Elec. Rev.*, New York, 1899, vol. xxxv. p. 375) applied in Russia to the manufacture of alcohol, by a series of chemical reactions starting from the production of acetylene by the action of water upon calcium carbide. The production of *ozone* in small quantities during electrolysis, and by the so-called silent discharge, has long been known, and the Siemens induction tube has been developed for use industrially. The Siemens and Halske ozonizer, in form somewhat resembling the old laboratory instrument, is largely used in Germany; working with an alternating current transformed up to 6500 volts, it has been found to give 280 grains or more of ozone per e.h.p. hour. E. Andreoli (whose first British ozone patent was No. 17,426 of 1891) uses flat aluminium plates and points, and working with an alternating current of 3000 volts is said to have obtained 1440 grains per e.h.p. hour. Yarnold's process, using corrugated glass plates coated on one side with gold or other metal leaf, is stated to have yielded as much as 2700 grains per e.h.p. hour. The ozone so prepared has numerous uses, as, for example, in bleaching oils, waxes, fabrics, &c., sterilizing drinking-water, maturing wines, cleansing foul beer-casks, oxidizing oil, and in the manufacture of vanillin. 210

For further information the following books, among others, may be consulted:—Haber, *Grundriss der technischen Elektrochemie* (München, 1898); Borchers and M'Millan, *Electric Smelting and Refining* (London, 1904); E.D. Peters, *Principles of Copper Smelting* (New York, 1907); F. Peters, *Angewandte Elektrochemie*, vols. ii. and iii. (Leipzig, 1900); Gore, *The Art of Electrolytic Separation of Metals* (London, 1890); Blount, *Practical Electro-Chemistry* (London, 1906); G. Langbein, *Vollständiges Handbuch der galvanischen Metall-Niederschläge* (Leipzig, 1903), Eng. trans. by W.T. Brannt (1909); A. Watt, *Electro-Plating and Electro-Refining of Metals* (London, 1902); W.H. Wahl, *Practical Guide to the Gold and Silver Electroplater, &c.* (Philadelphia, 1883); Wilson, *Stereotyping and Electrotyping* (London); Lunge, *Sulphuric Acid and Alkali*, vol. iii. (London, 1909). Also papers in various technical periodicals. The industrial aspect is treated in a Gartside Report, *Some Electro-Chemical Centres* (Manchester, 1908), by J.N. Pring.

(W. G. M.)

ELECTROCUTION (an anomalous derivative from "electro-execution"; syn. "electrothanasia"), the popular name, invented in America, for the infliction of the death penalty on criminals (see Capital Punishment) by passing through the body of the condemned a sufficient current of electricity to cause death. The method was first adopted by the state of New York, a law making this method obligatory having been passed and approved by the governor on the 4th of June 1888. The law provides that there shall be present, in addition to the warden, two physicians, twelve reputable citizens of full age, seven deputy sheriffs, and such ministers, priests or clergymen, not exceeding two, as the criminal may request. A post-mortem examination of the body of the convict is required, and the body, unless claimed by relatives, is interred in the prison cemetery with a sufficient quantity of quicklime to consume it. The law became effective in New York on the 1st of January 1889. The first criminal to be executed by electricity was William Kemmler, on the 6th of August 1890, at Auburn prison. The validity of the New York law had previously been attacked in regard to this case (*Re Kemmler*, 1889; 136 U.S. 436), as providing "a cruel and unusual punishment" and therefore being contrary to the Constitution; but it was sustained in the state courts and finally in the Federal courts. By 1906 about one hundred and fifteen murderers had been successfully executed by electricity in New York state in Sing Sing, Auburn and Dannemora prisons. The method has also been adopted by the states of Ohio (1896), Massachusetts (1898), New Jersey (1906), Virginia (1908) and North Carolina (1910).

The apparatus consists of a stationary engine, an alternating dynamo capable of generating a current at a pressure of 2000 volts, a "death-chair" with adjustable head-rest, binding straps and adjustable electrodes devised by E.F. Davis, the state electrician of New York. The voltmeter, ammeter and switch-board controlling the current are located in the execution-room; the dynamo-room is communicated with by electric signals. Before each execution the entire apparatus is thoroughly tested. When everything is in readiness the criminal is brought in and seats himself in the death-chair. His head, chest, arms and legs are secured by broad straps; one electrode thoroughly moistened with salt-solution is affixed to the head, and another to the calf of one leg, both electrodes being moulded so as to secure good contact. The application of the current is usually as follows: the contact is made with a high voltage (1700-1800 volts) for 5 to 7 seconds, reduced to 200 volts until a half-minute has elapsed; raised to high voltage for 3 to 5 seconds, again reduced to low voltage for 3 to 5 seconds, again reduced to a low voltage until one minute has elapsed, when it is again raised to the high voltage for a few seconds and the contact broken. The ammeter usually shows that from 7 to 10 amperes pass through the criminal's body. A second or even a third brief contact is sometimes made, partly as a precautionary measure, but rather the more completely to abolish reflexes in the dead body. Calculations have shown that by this method of execution from 7 to 10 h. p. of energy are liberated in the criminal's body. The time consumed by the strapping-in process is usually about 45 seconds, and the first contact is made about 70 seconds after the criminal has entered the death-chamber.

When properly performed the effect is painless and instantaneous death. The mechanism of life, circulation and respiration cease with the first contact. Consciousness is blotted out instantly, and the prolonged application of the current ensures permanent derangement of the vital functions beyond recovery. Occasionally the drying of the sponges through undue generation of heat causes desquamation or superficial blistering of the skin at the site of the electrodes. Post-mortem discoloration, or post-mortem lividity, often appears during the first contact. The pupils of the eyes dilate instantly and remain dilated after death.

The post-mortem examination of “electrocuted” criminals reveals a number of interesting phenomena. The temperature of the body rises promptly after death to a very high point. At the site of the leg electrode a temperature of over 128° F. was registered within fifteen minutes in many cases. After the removal of the brain the temperature recorded in the spinal canal was often over 120° F. The development of this high temperature is to be regarded as resulting from the active metabolism of tissues not (somatically) dead within a body where all vital mechanisms have been abolished, there being no circulation to carry off the generated heat. The heart, at first flaccid when exposed soon after death, gradually contracts and assumes a tetanized condition; it empties itself of all blood and takes the form of a heart in systole. The lungs are usually devoid of blood and weigh only 7 or 8 ounces (avoird.) each. The blood is profoundly altered biochemically; it is of a very dark colour and it rarely coagulates.

(E. A. S.*)

ELECTROKINETICS, that part of electrical science which is concerned with the properties of electric currents.

Classification of Electric Currents.—Electric currents are classified into (*a*) conduction currents, (*b*) convection currents, (*c*) displacement or dielectric currents. In the case of conduction currents electricity flows or moves through a stationary material body called the conductor. In convection currents electricity is carried from place to place with and on moving material bodies or particles. In dielectric currents there is no continued movement of electricity, but merely a limited displacement through or in the mass of an insulator or dielectric. The path in which an electric current exists is called an electric circuit, and may consist wholly of a conducting body, or partly of a conductor and insulator or dielectric, or wholly of a dielectric. In cases in which the three classes of currents are present together the true current is the sum of each separately. In the case of conduction currents the circuit consists of a conductor immersed in a non-conductor, and may take the form of a thin wire or cylinder, a sheet, surface or solid. Electric conduction currents may take place in space of one, two or three dimensions, but for the most part the circuits we have to consider consist of thin 211
cylindrical wires or tubes of conducting material surrounded with an insulator; hence the case which generally presents itself is that of electric flow in space of one dimension. Self-closed electric currents taking place in a sheet of conductor are called “eddy currents.”

Although in ordinary language the current is said to flow in the conductor, yet according to modern views the real pathway of the energy transmitted is the surrounding dielectric, and the so-called conductor or wire merely guides the transmission of energy in a certain direction. The presence of an electric current is recognized by three qualities or powers: (1) by the production of a magnetic field, (2) in the case of conduction currents, by the production of heat in the conductor, and (3) if the conductor is an electrolyte and the current unidirectional, by the occurrence of chemical decomposition in it. An electric current may also be regarded as the result of a movement of electricity across each section of the circuit, and is then measured by the quantity conveyed per unit of time. Hence if dq is the quantity of electricity which flows across any section of the conductor in the element of time dt, the current $i = dq/dt$.

Electric currents may be also classified as constant or variable and as unidirectional or “direct,” that is flowing always in the same direction, or “alternating,” that is reversing their direction at regular intervals. In the last case the variation of current may follow any particular law. It is called a “periodic current” if the cycle of current values is repeated during a certain time called the periodic time, during which the current reaches a certain maximum value, first in one direction and then in the opposite, and in the intervals between has a zero value at certain instants. The frequency of the periodic current is the number of periods or cycles in one second, and alternating currents are described as low frequency or high frequency, in the latter case having some thousands of periods per second. A periodic current may be represented either by a wave diagram, or by a polar diagram.[1] In the first case we take a straight line to represent the uniform flow of time, and at small equidistant intervals set up perpendiculars above or below the time axis, representing to scale the current at that instant in one direction or the other; the extremities of these ordinates then define a wavy curve which is called the wave form of the current (fig. 1). It is obvious that this curve can only be a single valued curve. In one particular and important case the form of the current curve is a simple harmonic curve or simple sine curve. If T represents the periodic time in which the cycle of current values takes place, whilst n is the frequency or number of periods per second and p stands for $2\pi n$, and i is the value of the current at any instant t, and I its maximum value, then in this case we have $i = I \sin pt$. Such a current is called a “sine current” or simple periodic current.

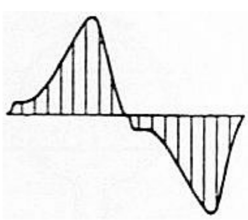

Fig. 1.

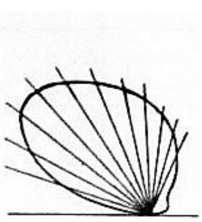

Fig. 2.

In a polar diagram (fig. 2) a number of radial lines are drawn from a point at small equiangular intervals, and on these lines are set off lengths proportional to the current value of a periodic current at corresponding intervals during one complete period represented by four right angles. The extremities of these radii delineate a polar curve. The polar form of a simple sine current is obviously a circle drawn through the origin. As a consequence of Fourier’s theorem it follows that any periodic curve having any wave form can be imitated by the superposition of simple sine currents differing in maximum value and in phase.

Definitions of Unit Electric Current.—In electrokinetic investigations we are most commonly limited to the cases of unidirectional continuous and constant currents (C.C. or D.C.), or of simple periodic currents, or alternating currents of sine form (A.C.). A continuous electric current is measured either by the magnetic effect it produces at some point outside its circuit, or by the amount of electrochemical decomposition it can perform in a given time on a selected standard electrolyte. Limiting our consideration to the case of linear currents or currents flowing in thin cylindrical wires, a definition may be given in the first place of the unit electric current in the centimetre, gramme, second (C.G.S.) of electromagnetic measurement (see Units, Physical). H.C. Oersted discovered in 1820 that a straight wire conveying an electric current is surrounded by a magnetic field the lines of which are self-closed lines embracing the electric circuit (see Electricity and Electromagnetism). The unit current in the electromagnetic system of measurement is defined as the current which, flowing in a thin wire bent into the form of a circle of one centimetre in radius, creates a magnetic field having a strength of 2π units at the centre of the circle, and therefore would exert a mechanical force of 2π dynes on a unit magnetic pole placed at that point (see Magnetism). Since the length of the circumference of the circle of unit radius is 2π units, this is equivalent to stating that the unit current on the electromagnetic C.G.S. system is a current such that unit length acts on unit magnetic pole with a unit force at a unit of distance. Another definition, called the electrostatic unit of current, is as follows: Let any conductor be charged with electricity and discharged through a thin wire at such a rate that one electrostatic unit of quantity (see Electrostatics) flows past any section of the wire in one unit of time. The electromagnetic unit of current defined as above is 3×10^{10} times larger than the electrostatic unit.

In the selection of a practical unit of current it was considered that the electromagnetic unit was too large for most purposes, whilst the electrostatic unit was too small; hence a practical unit of current called 1 ampere was selected, intended originally to be $1/10$ of the absolute electromagnetic C.G.S. unit of current as above defined. The practical unit of current, called the international ampere, is, however, legally defined at the present time as the continuous unidirectional current which when flowing through a neutral solution of silver nitrate deposits in one second on the cathode or negative pole 0.001118 of a gramme of silver. There is reason to believe that the international unit is smaller by about one part in a thousand, or perhaps by one part in 800, than the theoretical ampere defined as $1/10$ part of the absolute electromagnetic unit. A periodic or alternating current is said to have a value of 1 ampere if when passed through a fine wire it produces in the same time the same heat as a unidirectional continuous current of 1 ampere as above electrochemically defined. In the case of a simple periodic alternating current having a simple sine wave form, the maximum value is equal to that of the equiheating continuous current multiplied by $\sqrt{2}$. This equiheating continuous current is called the effective or root-mean-square (R.M.S.) value of the alternating one.

Resistance.—A current flows in a circuit in virtue of an electromotive force (E.M.F.), and the numerical relation between the current and E.M.F. is determined by three qualities of the circuit called respectively, its resistance (R), inductance (L), and capacity (C). If we limit our consideration to the case of continuous unidirectional conduction currents, then the relation between current and E.M.F. is defined by Ohm’s law, which states that the numerical value of the current is obtained as the quotient of the electromotive force by a certain constant of the circuit called its resistance, which is a function of the geometrical form of the circuit, of its nature, *i.e.* material, and of its temperature, but is independent of the electromotive force or current. The resistance (R) is measured in units called ohms and the electromotive force in volts (V); hence for a continuous current the value of the current in amperes (A) is obtained as the quotient of the 212
electromotive force acting in the circuit reckoned in volts by the resistance in ohms, or A = V/R. Ohm established his law by a course of reasoning which was similar to that on which J.B.J. Fourier based his investigations on the uniform motion of heat in a conductor. As a matter of fact, however, Ohm’s law merely states the direct proportionality of steady current to steady electromotive force in a circuit, and asserts that this ratio is governed by the numerical value of a quality of the conductor, called its resistance, which is independent of the current, provided that a correction is made for the change of temperature produced by the current. Our belief, however, in its universality and accuracy rests upon the close agreement between deductions made from it and observational results, and although it is not derivable from any more fundamental principle, it is yet one of the most certainly ascertained laws of electrokinetics.

Ohm’s law not only applies to the circuit as a whole but to any part of it, and provided the part selected does not contain a source of electromotive force it may be expressed as follows:—The difference of potential (P.D.) between any two points of a circuit including a resistance R, but not including any source of electromotive force, is proportional to the product of the resistance and the current i in the element, provided the conductor remains at the same temperature and the current is constant and unidirectional. If the current is varying we have, however, to take into account the electromotive force (E.M.F.) produced by this variation, and the product Ri is then equal to the difference between the observed P.D. and induced E.M.F.

We may otherwise define the resistance of a circuit by saying that it is that physical quality of it in virtue of which energy is dissipated as heat in the circuit when a current flows through it. The power communicated to any electric circuit when a current i is created in it by a continuous unidirectional electromotive force E is equal to Ei, and the energy dissipated as heat in that circuit by the conductor in a small interval of time dt is measured by Ei dt. Since by Ohm’s law E = Ri, where R is the resistance of the circuit, it follows that the energy dissipated as heat per unit of time in any circuit is numerically represented by Ri^2, and therefore the resistance is measured by the heat produced per unit of current, provided the current is unvarying.

Inductance.—As soon as we turn our attention, however, to alternating or periodic currents we find ourselves compelled to take into account another quality of the circuit, called its “inductance.” This may be defined as that quality in virtue of which energy is stored up in connexion with the circuit in a magnetic form. It can be experimentally shown that a current cannot be created instantaneously in a circuit by any finite electromotive force, and that when once created it cannot be annihilated instantaneously. The circuit possesses a quality analogous to the inertia of matter. If a current i is flowing in a circuit at any moment, the energy stored up in connexion with the circuit is measured by $\frac{1}{2}Li^2$, where L, the inductance of the circuit, is related to the current in the same manner as the quantity called the mass of a body is related to its velocity in the expression for the ordinary kinetic energy, viz. $\frac{1}{2}Mv^2$. The rate at which this conserved energy varies with the current is called the "electrokinetic momentum" of this circuit (= Li). Physically interpreted this quantity signifies the number of lines of magnetic flux due to the current itself which are self-linked with its own circuit.

Magnetic Force and Electric Currents.—In the case of every circuit conveying a current there is a certain magnetic force (see MAGNETISM) at external points which can in some instances be calculated. Laplace proved that the magnetic force due to an element of length dS of a circuit conveying a current I at a point P at a distance r from the element is expressed by $IdS \sin \theta / r^2$, where θ is the angle between the direction of the current element and that drawn between the element and the point. This force is in a direction perpendicular to the radius vector and to the plane containing it and the element of current. Hence the determination of the magnetic force due to any circuit is reduced to a summation of the effects due to all the elements of length. For instance, the magnetic force at the centre of a circular circuit of radius r carrying a steady current I is $2\pi I/r$, since all elements are at the same distance from the centre. In the same manner, if we take a point in a line at right angles to the plane of the circle through its centre and at a distance d, the magnetic force along this line is expressed by $2\pi r^2 I / (r^2 + d^2)^{3/2}$. Another important case is that of an infinitely long straight current. By summing up the magnetic force due to each element at any point P outside the continuous straight current I, and at a distance d from it, we can show that it is equal to 2I/d or is inversely proportional to the distance of the point from the wire. In the above formula the current I is measured in absolute electromagnetic units. If we reckon the current in amperes A, then I = A/10.

It is possible to make use of this last formula, coupled with an experimental fact, to prove that the magnetic force due to an element of current varies inversely as the square of the distance. If a flat circular disk is suspended so as to be free to rotate round a straight current which passes through its centre, and two bar magnets are placed on it with their axes in line with the current, it is found that the disk has no tendency to rotate round the current. This proves that the force on each magnetic pole is inversely as its distance from the current. But it can be shown that this law of action of the whole infinitely long straight current is a mathematical consequence of the fact that each element of the current exerts a magnetic force which varies inversely as the square of the distance. If the current flows N times round the circuit instead of once, we have to insert NA/10 in place of I in all the above formulae. The quantity NA is called the "ampere-turns" on the circuit, and it is seen that the magnetic field at any point outside a circuit is proportional to the ampere-turns on it and to a function of its geometrical form and the distance of the point.

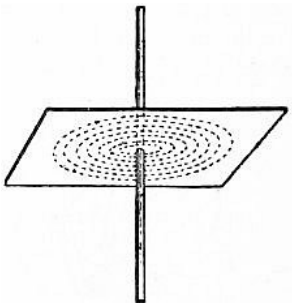

FIG. 3.

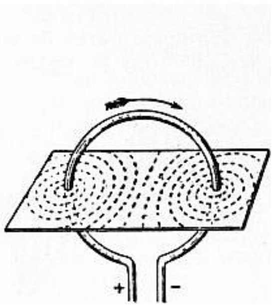

FIG. 4.

There is therefore a distribution of magnetic force in the field of every current-carrying conductor which can be delineated by lines of magnetic force and rendered visible to the eye by iron filings (see Magnetism). If a copper wire is passed vertically through a hole in a card on which iron filings are sprinkled, and a strong electric current is sent through the circuit, the filings arrange themselves in concentric circular lines making visible the paths of the lines of magnetic force (fig. 3). In the same manner, by passing a circular wire through a card and sending a strong current through the wire we can employ iron filings to delineate for us the form of the lines of magnetic force (fig. 4). In all cases a magnetic pole of strength M, placed in the field of an electric current, is urged along the lines of force with a mechanical force equal to MH, where H is the magnetic force. If then we carry a unit magnetic pole against the direction in which it would naturally move we do *work*. The lines of magnetic force embracing a current-carrying conductor are always loops or endless lines.

The work done in carrying a unit magnetic pole once round a circuit conveying a current is called the "line integral of magnetic force" along that path. If, for instance, we carry a unit pole in a circular path of radius r once round an infinitely long straight filamentary current I, the line integral is $4\pi I$. It is easy to prove that this is a general law, and that if we have any currents flowing in a conductor the line integral of magnetic force taken once round a path linked with the current circuit is 4π times the total current flowing through the circuit. Let us apply this to the case of an endless solenoid. If a copper wire insulated or covered with cotton or silk is twisted round a thin rod so as to make a close spiral, this forms a "solenoid," and if the solenoid is bent round so that its two ends come together we have an endless solenoid. Consider such a solenoid of mean length l and N turns of wire. If it is made endless, the magnetic force H is the same everywhere along the central axis and the line integral along the axis is Hl. If the current is denoted by I, then NI is the total current, and accordingly $4\pi NI = Hl$, or $H = 4\pi NI/l$. For a thin endless solenoid the axial magnetic force is therefore 4π times the current-turns per unit of length. This holds good also for a long straight solenoid provided its length is large compared with its diameter. It can be shown that if insulated wire is wound round a sphere, the turns being all parallel to lines of latitude, the magnetic force in the interior is constant and the lines of force therefore parallel. The magnetic force at a point outside a conductor conveying a current can by various means be measured or compared with some other standard magnetic forces, and it becomes then a means of measuring the current. Instruments called galvanometers and ammeters for the most part operate on this principle.

Thermal Effects of Currents.—J.P. Joule proved that the heat produced by a constant current in a given time in a wire having a constant resistance is proportional to the square of the strength of the current. This is known as Joule's law, and it follows, as already shown, as an immediate consequence of Ohm's law and the fact that the power dissipated electrically in a conductor, when an electromotive force E is applied to its extremities, producing thereby a current I in it, is equal to EI.

If the current is alternating or periodic, the heat produced in any time T is obtained by taking the sum at equidistant intervals of time of all the values of the quantities Ri^2dt, where dt represents a small interval of time and i is the current at that instant. The quantity $T^{-1}\int_0^T i^2dt$ is called the mean-square-value of the variable current, i being the instantaneous value of the current, that is, its value at a particular instant or during a very small interval of time dt. The square root of the above quantity, or

$$\left[T^{-1}\int_0^T i^2dt \right]^{1/2},$$

is called the root-mean-square-value, or the effective value of the current, and is denoted by the letters R.M.S.

Currents have equal heat-producing power in conductors of identical resistance when they have the same R.M.S. values. Hence periodic or alternating currents can be measured as regards their R.M.S. value by ascertaining the continuous current which produces in the same time the same heat in the same conductor as the periodic current considered. Current measuring instruments depending on this fact, called hot-wire ammeters, are in common use, especially for measuring alternating currents. The maximum value of the periodic current can only be determined from the R.M.S. value when we know the wave form of the current. The thermal effects of electric currents in conductors are dependent upon the production of a state of equilibrium between the heat produced electrically in the wire and the causes operative in removing it. If an ordinary round wire is heated by a current it loses heat, (1) by radiation, (2) by air convection or cooling, and (3) by conduction of heat out of the ends of the wire. Generally speaking, the greater part of the heat removal is effected by radiation and convection.

If a round sectioned metallic wire of uniform diameter d and length l made of a material of resistivity ρ has a current of A amperes passed through it, the heat in watts produced in any time t seconds is represented by the value of $4A^2\rho lt / 10^9\pi d^2$, where d and l must be measured in centimetres and ρ in absolute C.G.S. electromagnetic units. The factor 10^9 enters because one ohm is 10^9 absolute electromagnetic C.G.S. units (see Units, Physical). If the wire has an emissivity e, by which is meant that e units of heat reckoned in joules or watt-seconds are radiated per second from unit of surface, then the power removed by radiation in the time t is expressed by πdlet. Hence when thermal equilibrium is established we have $4A^2\rho lt / 10^9\pi d^2 = \pi dlet$, or $A^2 = 10^9\pi^2ed^3 / 4\rho$. If the diameter of the wire is reckoned in mils (1 mil = .001 in.), and if we take e to have a value 0.1, an emissivity which will generally bring the wire to about 60° C., we can put the above formula in the following forms for circular sectioned copper, iron or platinoid wires, viz.

$$A = \sqrt{d^3 / 500} \text{ for copper wires}$$
$$A = \sqrt{d^3 / 4000} \text{ for iron wires}$$
$$A = \sqrt{d^3 / 5000} \text{ for platinoid wires.}$$

These expressions give the ampere value of the current which will bring bare, straight or loosely coiled wires of d mils in diameter to about 60° C. when the steady state of temperature is reached. Thus, for instance, a bare straight copper wire 50 mils in diameter (= 0.05 in.) will be brought to a steady temperature of about 60° C. if a current of $\sqrt{50^3/500} = \sqrt{250} = 16$ amperes (nearly) is passed through it, whilst a current of $\sqrt{25} = 5$ amperes would bring a platinoid wire to about the same temperature.

A wire has therefore a certain safe current-carrying capacity which is determined by its specific resistance and emissivity, the latter being fixed by its form, surface and surroundings. The emissivity increases with the temperature, else no state of thermal equilibrium could be reached. It has been found experimentally that whilst for fairly thick wires from 8 to 60 mils in diameter the safe current varies approximately as the 1.5th power of the diameter, for fine wires of 1 to 3 mils it varies more nearly as the diameter.

Action of one Current on Another.—The investigations of Ampère in connexion with electric currents are of fundamental importance in electrokinetics. Starting from the discovery of Oersted, Ampère made known the correlative fact that not only is there a mechanical action between a current and a magnet, but that two conductors conveying electric currents exert mechanical forces on each other. Ampère devised ingenious methods of making one portion of a circuit movable so that he might observe effects of attraction or repulsion between this circuit and some other fixed current. He employed for this purpose an astatic circuit B, consisting of a wire bent into a double rectangle round which a current flowed first in one and then in the opposite direction (fig. 5). In this way the circuit was removed from the action of the earth's magnetic field, and yet one portion of it could be submitted to the action of any other circuit C. The astatic circuit was pivoted by suspending it in mercury cups q, p, one of which was in electrical connexion with the tubular support A, and the other with a strong insulated wire passing up it.

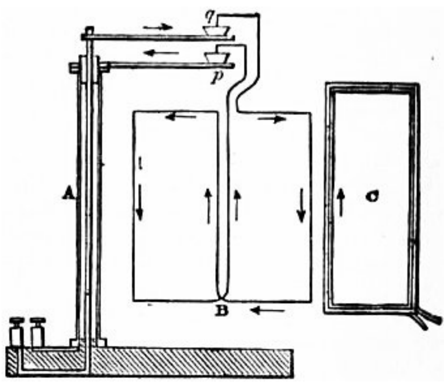

Fig. 5.

Ampère devised certain crucial experiments, and the theory deduced from them is based upon four facts and one assumption.[2] He showed (1) that wire conveying a current bent back on itself produced no action upon a proximate portion of a movable astatic circuit; (2) that if the return wire was bent zig-zag but close to the outgoing straight wire the circuit produced no action on the movable one, showing that the effect of an element of the circuit was proportional to its projected length; (3) that a closed circuit cannot cause motion in an element of another circuit free to move in the direction of its length; and (4) that the action of two circuits on one and the same movable circuit was null if one of the two fixed circuits was n times greater than the other but n times further removed from the movable circuit. From this last experiment by an ingenious line of reasoning he proved that the action of an element of current on another element of current varies inversely as a square of their distance. These experiments enabled him to construct a mathematical expression of the law of action between two elements of conductors conveying currents. They also enabled him to prove that an element of current may be resolved like a force into components in different directions, also that the force produced by any element of the circuit on an element of any other circuit was perpendicular to the line joining the elements and inversely as the square of their distance. Also he showed that this force was an attraction if the currents in the elements were in the same direction, but a repulsion if they were in opposite directions. From these experiments and deductions from them he built up a complete formula for the action of one element of a current of length dS of one conductor conveying a current I upon 214 another element dS′ of another circuit conveying another current I′ the elements being at a distance apart equal to r.

If θ and θ′ are the angles the elements make with the line joining them, and φ the angle they make with one another, then Ampère’s expression for the mechanical force f the elements exert on one another is

$$f = 2II'r^{-2} \{\cos\varphi - \tfrac{3}{2}\cos\theta\cos\theta'\}\, dSdS'.$$

This law, together with that of Laplace already mentioned, viz. that the magnetic force due to an element of length dS of a current I at a distance r, the element making an angle θ with the radius vector o is $IdS \sin\theta/r^2$, constitute the fundamental laws of electrokinetics.

Ampère applied these with great mathematical skill to elucidate the mechanical actions of currents on each other, and experimentally confirmed the following deductions: (1) Currents in parallel circuits flowing in the same direction attract each other, but if in opposite directions repel each other. (2) Currents in wires meeting at an angle attract each other more into parallelism if both flow either to or from the angle, but repel each other more widely apart if they are in opposite directions. (3) A current in a small circular conductor exerts a magnetic force in its centre perpendicular to its plane and is in all respects equivalent to a magnetic shell or a thin circular disk of steel so magnetized that one face is a north pole and the other a south pole, the product of the area of the circuit and the current flowing in it determining the magnetic moment of the element. (4) A closely wound spiral current is equivalent as regards external magnetic force to a polar magnet, such a circuit being called a finite solenoid. (5) Two finite solenoid circuits act on each other like two polar magnets, exhibiting actions of attraction or repulsion between their ends.

Ampère’s theory was wholly built up on the assumption of action at a distance between elements of conductors conveying the electric currents. Faraday’s researches and the discovery of the fact that the insulating medium is the real seat of the operations necessitates a change in the point of view from which we regard the facts discovered by Ampère. Maxwell showed that in any field of magnetic force there is a tension along the lines of force and a pressure at right angles to them; in other words, lines of magnetic force are like stretched elastic threads which tend to contract.[3] If, therefore, two conductors lie parallel and have currents in them in the same direction they are impressed by a certain number of lines of magnetic force which pass round the two conductors, and it is the tendency of these to contract which draws the circuits together. If, however, the currents are in opposite directions then the lateral pressure of the similarly contracted lines of force between them pushes the conductors apart. Practical application of Ampère’s discoveries was made by W.E. Weber in inventing the electrodynamometer, and later Lord Kelvin devised ampere balances for the measurement of electric currents based on the attraction between coils conveying electric currents.

Induction of Electric Currents.—Faraday[4] in 1831 made the important discovery of the induction of electric currents (see Electricity). If two conductors are placed parallel to each other, and a current in one of them, called the primary, started or stopped or changed in strength, every such alteration causes a transitory current to appear in the other circuit, called the secondary. This is due to the fact that as the primary current increases or decreases, its own embracing magnetic field alters, and lines of magnetic force are added to or subtracted from its fields. These lines do not appear instantly in their place at a distance, but are propagated out from the wire with a velocity equal to that of light; hence in their outward progress they cut through the secondary circuit, just as ripples made on the surface of water in a lake by throwing a stone on to it expand and cut through a stick held vertically in the water at a distance from the place of origin of the ripples. Faraday confirmed this view of the phenomena by proving that the mere motion of a wire transversely to the lines of magnetic force of a permanent magnet gave rise to an induced electromotive force in the wire. He embraced all the facts in the single statement that if there be any circuit which by movement in a magnetic field, or by the creation or change in magnetic fields round it, experiences a change in the number of lines of force linked with it, then an electromotive force is set up in that circuit which is proportional at any instant to the rate at which the total magnetic flux linked with it is changing. Hence if Z represents the total number of lines of magnetic force linked with a circuit of N turns, then $-N\,(dZ/dt)$ represents the electromotive force set up in that circuit. The operation of the induction coil (*q.v.*) and the transformer (*q.v.*) are based on this discovery. Faraday also found that if a copper disk A (fig. 6) is rotated between the poles of a magnet NO so that the disk moves with its plane perpendicular to the lines of magnetic force of the field, it has created in it an electromotive force directed from the centre to the edge or vice versa. The action of the dynamo (*q.v.*) depends on similar processes, viz. the cutting of the lines of magnetic force of a constant field produced by certain magnets by certain moving conductors called armature bars or coils in which an electromotive force is thereby created.

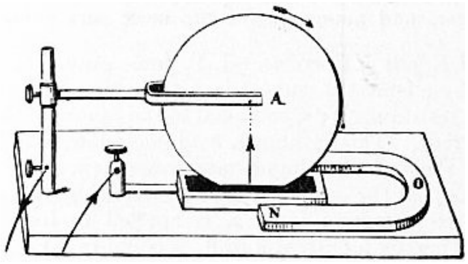

FIG 6.

In 1834 H.F.E. Lenz enunciated a law which connects together the mechanical actions between electric circuits discovered by Ampère and the induction of electric currents discovered by Faraday. It is as follows: If a constant current flows in a primary circuit P, and if by motion of P a secondary current is created in a neighbouring circuit S, the direction of the secondary current will be such as to oppose the relative motion of the circuits. Starting from this, F.E. Neumann founded a mathematical theory of induced currents, discovering a quantity M, called the "potential of one circuit on another," or generally their "coefficient of mutual inductance." Mathematically M is obtained by taking the sum of all such quantities as $\iint dSdS' \cos\varphi/r$, where dS and dS′ are the elements of length of the two circuits, r is their distance, and φ is the angle which they make with one another; the summation or integration must be extended over every possible pair of elements. If we take pairs of elements in the same circuit, then Neumann's formula gives us the coefficient of self-induction of the circuit or the potential of the circuit on itself. For the results of such calculations on various forms of circuit the reader must be referred to special treatises.

H. von Helmholtz, and later on Lord Kelvin, showed that the facts of induction of electric currents discovered by Faraday could have been predicted from the electrodynamic actions discovered by Ampère assuming the principle of the conservation of energy. Helmholtz takes the case of a circuit of resistance R in which acts an electromotive force due to a battery or thermopile. Let a magnet be in the neighbourhood, and the potential of the magnet on the circuit be V, so that if a current I existed in the circuit the work done on the magnet in the time dt is I (dV/dt)dt. The source of electromotive force supplies in the time dt work equal to EIdt, and according to Joule's law energy is dissipated equal to RI^2dt. Hence, by the conservation of energy,

$$EIdt = RI^2dt + I\,(dV/dt)\,dt.$$

If then E = 0, we have $I = -(dV/dt)\,/\,R$, or there will be a current due to an induced electromotive force expressed by $-dV/dt$. Hence if the magnet moves, it will create a current in the wire provided that such motion changes the potential of the magnet with respect to the circuit. This is the effect discovered by Faraday.[5]

Oscillatory Currents.—In considering the motion of electricity in conductors we find interesting phenomena connected with the discharge of a condenser or Leyden jar (*q.v.*). This problem was first mathematically treated by Lord Kelvin in 1853 (*Phil. Mag.*, 1853, 5, p. 292).

If a conductor of capacity C has its terminals connected by a wire of resistance R and inductance L, it becomes important to consider the subsequent motion of electricity in the wire. If Q is the quantity of 215 electricity in the condenser initially, and q that at any time t after completing the circuit, then the energy stored up in the condenser at that instant is $\tfrac{1}{2}q^2/C$, and the energy associated with the circuit is $\tfrac{1}{2}L\,(dq/dt)^2$, and the rate of dissipation of energy by resistance is $R\,(dq/dt)^2$, since $dq/dt = i$ is the discharge current. Hence we can construct an equation of energy which expresses the fact that at any instant the power given out by the condenser is partly stored in the circuit and partly dissipated as heat in it. Mathematically this is expressed as follows:—

$$-\frac{d}{dt}\left[\tfrac{1}{2}\frac{q^2}{C}\right] = \frac{d}{dt}\left[\tfrac{1}{2}L\left(\frac{dq}{dt}\right)^2\right] + R\left(\frac{dq}{dt}\right)^2$$

or

$$\frac{d^2q}{dt^2} + \frac{R}{L}\,\frac{dq}{dt} + \frac{1}{LC}\,q = 0.$$

The above equation has two solutions according as $R^2/4L^2$ is greater or less than 1/LC. In the first case the current i in the circuit can be expressed by the equation

$$i = Q\,\frac{\alpha^2+\beta^2}{2\beta}\,e^{-\alpha t}\,(e^{\beta t} - e^{-\beta t}),$$

where $\alpha = R/2L$, $\beta = \sqrt{(R^2/4L^2 - 1/LC)}$, Q is the value of q when t = 0, and e is the base of Napierian logarithms; and in the second case by the equation

$$i = Q\,\frac{\alpha^2+\beta^2}{\beta}\,e^{-\alpha t}\sin\beta t$$

where

$$\alpha = R/2L, \text{ and } \beta = \sqrt{\frac{1}{LC} - \frac{R^2}{4L^2}}\,.$$

These expressions show that in the first case the discharge current of the jar is always in the same direction and is a transient unidirectional current. In the second case, however, the current is an oscillatory current gradually decreasing in amplitude, the frequency n of the oscillation being given by the expression

$$n = \frac{1}{2\pi}\sqrt{\frac{1}{LC} - \frac{R^2}{4L^2}}\,.$$

In those cases in which the resistance of the discharge circuit is very small, the expression for the frequency n and for the time period of oscillation R take the simple forms $n = 1, 2\pi\sqrt{LC}$, or $T = 1/n = 2\pi\sqrt{LC}$.

The above investigation shows that if we construct a circuit consisting of a condenser and inductance placed in series with one another, such circuit has a natural electrical time period of its own in which the electrical charge in it oscillates if disturbed. It may therefore be compared with a pendulum of any kind which when displaced oscillates with a time period depending on its inertia and on its restoring force.

The study of these electrical oscillations received a great impetus after H.R. Hertz showed that when taking place in electric circuits of a certain kind they create electromagnetic waves (see Electric Waves) in the dielectric surrounding the oscillator, and an additional interest was given to them by their application to telegraphy. If a Leyden jar and a circuit of low resistance but some inductance in series with it are connected across the secondary spark gap of an induction coil, then when the coil is set in action we have a series of bright noisy sparks, each of which consists of a train of oscillatory electric discharges from the jar. The condenser becomes charged as the secondary electromotive force of the coil is created at each break of the primary current, and when the potential difference of the condenser coatings reaches a certain value determined by the spark-ball distance a discharge happens. This discharge, however, is not a single movement of electricity in one direction but an oscillatory motion with gradually decreasing amplitude. If the oscillatory spark is photographed on a revolving plate or a rapidly moving film, we have evidence in the photograph that such a spark consists of numerous intermittent sparks gradually becoming feebler. As the coil continues to operate, these trains of electric discharges take place at regular intervals. We can cause a train of electric oscillations in one circuit to induce similar oscillations in a neighbouring circuit, and thus construct an oscillation transformer or high frequency induction coil.

Alternating Currents.—The study of alternating currents of electricity began to attract great attention towards the end of the 19th century by reason of their application in electrotechnics and especially to the transmission of power. A circuit in which a simple periodic alternating current flows is called a single phase circuit. The important difference between such a form of current flow and steady current flow arises from the fact that if the circuit has inductance then the periodic electric current in it is not in step with the terminal potential difference or electromotive force acting in the circuit, but the current lags behind the electromotive force by a certain fraction of the periodic time called the “phase difference.” If two alternating currents having a fixed difference in phase flow in two connected separate but related circuits, the two are called a two-phase current. If three or more single-phase currents preserving a fixed difference of phase flow in various parts of a connected circuit, the whole taken together is called a polyphase current. Since an electric current is a vector quantity, that is, has direction as well as magnitude, it can most conveniently be represented by a line denoting its maximum value, and if the alternating current is a simple periodic current then the root-mean-square or effective value of the current is obtained by dividing the maximum value by $\sqrt{2}$. Accordingly when we have an electric circuit or circuits in which there are simple periodic currents we can draw a vector diagram, the lines of which represent the relative magnitudes and phase differences of these currents.

A vector can most conveniently be represented by a symbol such as $a + \iota b$, where a stands for any length of a units measured horizontally and b for a length b units measured vertically, and the symbol ι is a sign of perpendicularity, and equivalent analytically[6] to $\sqrt{-1}$. Accordingly if E represents the periodic electromotive force (maximum value) acting in a circuit of resistance R and inductance L and frequency n, and if the current considered as a vector is represented by I, it is easy to show that a vector equation exists between these quantities as follows:—

$$E = RI + \iota 2\pi nLI.$$

Since the absolute magnitude of a vector $a + \iota b$ is $\sqrt{(a^2 + b^2)}$, it follows that considering merely magnitudes of current and electromotive force and denoting them by symbols (E) (I), we have the following equation connecting (I) and (E):—

$$(I) = (E) / \sqrt{R^2 + p^2L^2},$$

where p stands for $2\pi n$. If the above equation is compared with the symbolic expression of Ohm’s law, it will be seen that the quantity $\sqrt{(R^2 + p^2L^2)}$ takes the place of resistance R in the expression of Ohm. This quantity $\sqrt{(R^2 + p^2L^2)}$ is called the “impedance” of the alternating circuit. The quantity pL is called the “reactance” of the alternating circuit, and it is therefore obvious that the current in such a circuit lags behind the electromotive force by an angle, called the angle of lag, the tangent of which is pL/R.

Currents in Networks of Conductors.—In dealing with problems connected with electric currents we have to consider the laws which govern the flow of currents in linear conductors (wires), in plane conductors (sheets), and throughout the mass of a material conductor.[7] In the first case consider the collocation of a number of linear conductors, such as rods or wires of metal, joined at their ends to form a network of conductors. The network consists of a number of conductors joining certain points and forming meshes. In each conductor a current may exist, and along each conductor there is a fall of potential, or an active electromotive force may be acting in it. Each conductor has a certain resistance. To find the current in each conductor when the individual resistances and electromotive forces are given, proceed as follows:—Consider any one mesh. The sum of all the electromotive forces which exist in the branches bounding that mesh must be equal to the sum of all the products of the resistances into the currents flowing along them, or $\Sigma(E) = \Sigma(C.R.)$. Hence if we consider each mesh as traversed by imaginary currents all circulating in the same direction, the real currents are the sums or differences of these imaginary cyclic currents in each branch. Hence we may assign to each mesh a cycle symbol x, y, z, &c., and form a cycle equation. Write down the cycle symbol for a mesh and prefix as coefficient the sum of all the resistances which bound that cycle, then subtract the cycle symbols of each adjacent cycle, each multiplied by the value of the bounding or common resistances, and equate this sum to the total electromotive force acting round the cycle. Thus if x y z are the cycle currents, and a b c the resistances bounding the mesh x, and b and c those

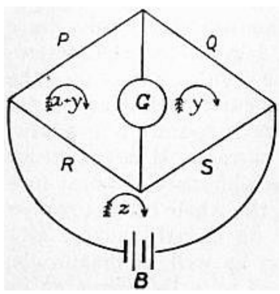

Fig. 7.

separating it from the meshes y and z, and E an electromotive force in the branch a, then we have formed the cycle equation x(a + b + c) − by − cz = E. For each mesh a similar equation may be formed. Hence we have as many linear equations as there are meshes, and we can obtain the solution for each cycle symbol, and therefore for the current in each branch. The solution giving the current in such branch of the network is therefore always in the form of the quotient of two determinants. The solution of the well-known problem of finding the current in the galvanometer circuit of the arrangement of linear conductors called Wheatstone's Bridge is thus easily obtained. For if we call the cycles (see fig. 7) (x + y), y and z, and the resistances P, Q, R, S, G and B, and if E be the electromotive force in the battery circuit, we have the cycle equations

$$(P + G + R)(x + y) - Gy - Rz = 0,$$
$$(Q + G + S)y - G(x + y) - Sz = 0,$$
$$(R + S + B)z - R(x + y) - Sy = E.$$

From these we can easily obtain the solution for (x + y) − y = x, which is the current through the galvanometer circuit in the form

$$x = E(PS - RQ)\,\Delta.$$

where Δ is a certain function of P, Q, R, S, B and G.

Currents in Sheets.—In the case of current flow in plane sheets, we have to consider certain points called sources at which the current flows into the sheet, and certain points called sinks at which it leaves. We may investigate, first, the simple case of one source and one sink in an infinite plane sheet of thickness δ and conductivity k. Take any point P in the plane at distances R and r from the source and sink respectively. The potential V at P is obviously given by

$$V = \frac{Q}{2\pi k\delta} \log_e \frac{r_1}{r_2},$$

where Q is the quantity of electricity supplied by the source per second. Hence the equation to the equipotential curve is $r_1 r_2$ = a constant.

If we take a point half-way between the sink and the source as the origin of a system of rectangular co-ordinates, and if the distance between sink and source is equal to p, and the line joining them is taken as the axis of x, then the equation to the equipotential line is

$$\frac{y^2 + (x + p)^2}{y^2 + (x - p)^2} = \text{a constant.}$$

This is the equation of a family of circles having the axis of y for a common radical axis, one set of circles surrounding the sink and another set of circles surrounding the source. In order to discover the form of the stream of current lines we have to determine the orthogonal trajectories to this family of coaxial circles. It is easy to show that the orthogonal trajectory of the system of circles is another system of circles all passing through the sink and the source, and as a corollary of this fact, that the electric resistance of a circular disk of uniform thickness is the same between any two points taken anywhere on its circumference as sink and source. These equipotential lines may be delineated experimentally by attaching the terminals of a battery or batteries to small wires which touch at various places a sheet of tinfoil. Two wires attached to a galvanometer may then be placed on the tinfoil, and one may be kept stationary and the other may be moved about, so that the galvanometer is not traversed by any current. The moving terminal then traces out an equipotential curve. If there are n sinks and sources in a plane conducting sheet, and if r, r′, r″ be the distances of any point from the sinks, and t, t′, t″ the distances of the sources, then

$$\frac{r\, r'\, r'' \ldots}{t\, t'\, t'' \ldots} = \text{a constant,}$$

is the equation to the equipotential lines. The orthogonal trajectories or stream lines have the equation

$$\Sigma(\theta - \theta') = \text{a constant,}$$

where θ and θ′ are the angles which the lines drawn from any point in the plane to the sink and corresponding source make with the line joining that sink and source. Generally it may be shown that if there are any number of sinks and sources in an infinite plane-conducting sheet, and if r, θ are the polar co-ordinates of any one, then the equation to the equipotential surfaces is given by the equation

$$\Sigma(A \log_e r) = \text{a constant,}$$

where A is a constant; and the equation to the stream of current lines is

$$\Sigma(\theta) = \text{a constant.}$$

In the case of electric flow in three dimensions the electric potential must satisfy Laplace's equation, and a solution is therefore found in the form Σ (A/r) = a constant, as the equation to an equipotential surface, where r is the distance of any point on that surface from a source or sink.

Convection Currents.—The subject of convection electric currents has risen to great importance in connexion with modern electrical investigations. The question whether a statically electrified body in motion creates a magnetic field is of fundamental importance. Experiments to settle it were first undertaken in the year 1876 by H.A. Rowland, at a suggestion of H. von Helmholtz.[8] After preliminary experiments, Rowland's first apparatus for testing this hypothesis was constructed, as follows:—An ebonite disk was covered with radial strips of gold-leaf and placed between two other metal plates which acted as screens. The disk was then charged with electricity and set in rapid rotation. It was found to affect a delicately suspended pair of astatic magnetic needles hung in proximity to the disk just as would, by Oersted's rule, a circular electric current coincident with the periphery of the disk. Hence the statically-charged but rotating disk becomes in effect a circular electric current.

The experiments were repeated and confirmed by W.C. Röntgen (*Wied. Ann.*, 1888, 35, p. 264; 1890, 40, p.

93) and by F. Himstedt (*Wied. Ann.*, 1889, 38, p. 560). Later V. Crémieu again repeated them and obtained negative results (*Com. rend.*, 1900, 130, p. 1544, and 131, pp. 578 and 797; 1901, 132, pp. 327 and 1108). They were again very carefully reconducted by H. Pender (*Phil. Mag.*, 1901, 2, p. 179) and by E.P. Adams (id. *ib.*, 285). Pender's work showed beyond any doubt that electric convection does produce a magnetic effect. Adams employed charged copper spheres rotating at a high speed in place of a disk, and was able to prove that the rotation of such spheres produced a magnetic field similar to that due to a circular current and agreeing numerically with the theoretical value. It has been shown by J.J. Thomson (*Phil. Mag.*, 1881, 2, p. 236) and O. Heaviside (*Electrical Papers*, vol. ii. p. 205) that an electrified sphere, moving with a velocity v and carrying a quantity of electricity q, should produce a magnetic force H, at a point at a distance ρ from the centre of the sphere, equal to $qv \sin \theta/\rho^2$, where θ is the angle between the direction of ρ and the motion of the sphere. Adams found the field produced by a known electric charge rotating at a known speed had a strength not very different from that predetermined by the above formula. An observation recorded by R.W. Wood (*Phil. Mag.*, 1902, 2, p. 659) provides a confirmatory fact. He noticed that if carbon-dioxide strongly compressed in a steel bottle is allowed to escape suddenly the cold produced solidifies some part of the gas, and the issuing jet is full of particles of carbon-dioxide snow. These by friction against the nozzle are electrified positively. Wood caused the jet of gas to pass through a glass tube 2.5 mm. in diameter, and found that these particles of electrified snow were blown through it with a velocity of 2000 ft. a second. Moreover, he found that a magnetic needle hung near the tube was deflected as if held near an electric current. Hence the positively electrified particles in motion in the tube create a magnetic field round it.

Nature of an Electric Current.—The question, What is an electric current? is involved in the larger question of the nature of electricity. Modern investigations have shown that negative electricity is identical with the electrons or corpuscles which are components of the chemical atom (see Matter and Electricity). Certain lines of argument lead to the conclusion that a solid conductor is not only composed of chemical atoms, but that there is a certain proportion of free electrons present in it, the electronic density or number per unit of volume being determined by the material, its temperature and other physical conditions. If any cause operates to add or remove electrons at one point there is an immediate diffusion of electrons to re-establish equilibrium, and this electronic movement constitutes an electric current. This hypothesis explains the reason for the identity between the laws of diffusion of matter, of heat and of electricity. Electromotive force is then any cause making or tending to make an inequality of electronic density in conductors, and may arise from differences of temperature, *i.e.* thermoelectromotive force (see Thermoelectricity), or from chemical 217
action when part of the circuit is an electrolytic conductor, or from the movement of lines of magnetic force across the conductor.

Bibliography.—For additional information the reader may be referred to the following books: M. Faraday, *Experimental Researches in Electricity* (3 vols., London, 1839, 1844, 1855); J. Clerk Maxwell, *Electricity and Magnetism* (2 vols., Oxford, 1892); W. Watson and S.H. Burbury, *Mathematical Theory of Electricity and Magnetism*, vol. ii. (Oxford, 1889); E. Mascart and J. Joubert, *A Treatise on Electricity and Magnetism* (2 vols., London, 1883); A. Hay, *Alternating Currents* (London, 1905); W.G. Rhodes, *An Elementary Treatise on Alternating Currents* (London, 1902); D.C. Jackson and J.P. Jackson, *Alternating Currents and Alternating Current Machinery* (1896, new ed. 1903); S.P. Thompson, *Polyphase Electric Currents* (London, 1900); *Dynamo-Electric Machinery*, vol. ii., "Alternating Currents" (London, 1905); E.E. Fournier d'Albe, *The Electron Theory* (London, 1906).

(J. A. F.)

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1 See J.A. Fleming, *The Alternate Current Transformer*, vol. i. p. 519.

2 See Maxwell, *Electricity and Magnetism*, vol. ii. chap. ii.

3 See Maxwell, *Electricity and Magnetism*, vol. ii. 642.

4 *Experimental Researches*, vol. i. ser. 1.

5 See Maxwell, *Electricity and Magnetism*, vol. ii. § 542, p. 178.

6 See W.G. Rhodes, *An Elementary Treatise on Alternating Currents* (London, 1902), chap. vii.

7 See J.A. Fleming, "Problems on the Distribution of Electric Currents in Networks of Conductors," *Phil. Mag.* (1885), or Proc. Phys. Soc. Lond. (1885), 7; also Maxwell, *Electricity and Magnetism* (2nd ed.), vol. i. p. 374, § 280, 282b.

8 See *Berl. Acad. Ber.*, 1876, p. 211; also H.A. Rowland and C.T. Hutchinson, "On the Electromagnetic Effect of Convection Currents," *Phil. Mag.*, 1889, 27, p. 445.

ELECTROLIER, a fixture, usually pendent from the ceiling, for holding electric lamps. The word is analogous to chandelier, from which indeed it was formed.

ELECTROLYSIS (formed from Gr. λύειν, to loosen). When the passage of an electric current through a substance is accompanied by definite chemical changes which are independent of the heating effects of the current, the process is known as *electrolysis*, and the substance is called an *electrolyte*. As an example we may take the case of a solution of a salt such as copper sulphate in water, through which an electric current is passed between copper plates. We shall then observe the following phenomena. (1) The bulk of the solution is unaltered, except that its temperature may be raised owing to the usual heating effect which is proportional to the square of the strength of the current. (2) The copper plate by which the current is said to enter the solution, *i.e.* the plate attached to the so-called positive terminal of the battery or other source of current, dissolves away, the copper going into solution as copper sulphate. (3) Copper is deposited on the surface of the other plate, being obtained from the solution. (4) Changes in concentration are produced in the neighbourhood of the two plates or electrodes. In the case we have chosen, the solution becomes stronger near the anode, or electrode at which the current enters, and weaker near the cathode, or electrode at which it leaves the solution. If, instead of using copper electrodes, we take plates of platinum, copper is still deposited on the cathode; but, instead of the anode dissolving, free sulphuric acid appears in the neighbouring solution, and oxygen gas is evolved at the surface of the platinum plate.

With other electrolytes similar phenomena appear, though the primary chemical changes may be masked by secondary actions. Thus, with a dilute solution of sulphuric acid and platinum electrodes, hydrogen gas is evolved at the cathode, while, as the result of a secondary action on the anode, sulphuric acid is there re-formed, and oxygen gas evolved. Again, with the solution of a salt such as sodium chloride, the sodium, which is primarily liberated at the cathode, decomposes the water and evolves hydrogen, while the chlorine may be evolved as such, may dissolve the anode, or may liberate oxygen from the water, according to the nature of the plate and the concentration of the solution.

Early History of Electrolysis.—Alessandro Volta of Pavia discovered the electric battery in the year 1800, and thus placed the means of maintaining a steady electric current in the hands of investigators, who, before that date, had been restricted to the study of the isolated electric charges given by frictional electric machines. Volta's cell consists essentially of two plates of different metals, such as zinc and copper, connected by an electrolyte such as a solution of salt or acid. Immediately on its discovery intense interest was aroused in the new invention, and the chemical effects of electric currents were speedily detected. W. Nicholson and Sir A. Carlisle found that hydrogen and oxygen were evolved at the surfaces of gold and platinum wires connected with the terminals of a battery and dipped in water. The volume of the hydrogen was about double that of the oxygen, and, since this is the ratio in which these elements are combined in water, it was concluded that the process consisted essentially in the decomposition of water. They also noticed that a similar kind of chemical action went on in the battery itself. Soon afterwards, William Cruickshank decomposed the magnesium, sodium and ammonium chlorides, and precipitated silver and copper from their solutions—an observation which led to the process of electroplating. He also found that the liquid round the anode became acid, and that round the cathode alkaline. In 1804 W. Hisinger and J.J. Berzelius stated that neutral salt solutions could be decomposed by electricity, the acid appearing at one pole and the metal at the other. This observation showed that nascent hydrogen was not, as had been supposed, the primary cause of the separation of metals from their solutions, but that the action consisted in a direct decomposition into metal and acid. During the earliest investigation of the subject it was thought that, since hydrogen and oxygen were usually evolved, the electrolysis of solutions of acids and alkalis was to be regarded as a direct decomposition of water. In 1806 Sir Humphry Davy proved that the formation of acid and alkali when water was electrolysed was due to saline impurities in the water. He had shown previously that decomposition of water could be effected although the two poles were placed in separate vessels connected by moistened threads. In 1807 he decomposed potash and soda, previously considered to be elements, by passing the current from a powerful battery through the moistened solids, and thus isolated the metals potassium and sodium.

The electromotive force of Volta's simple cell falls off rapidly when the cell is used, and this phenomenon was shown to be due to the accumulation at the metal plates of the products of chemical changes in the cell itself. This reverse electromotive force of polarization is produced in all electrolytes when the passage of the current changes the nature of the electrodes. In batteries which use acids as the electrolyte, a film of hydrogen tends to be deposited on the copper or platinum electrode; but, to obtain a constant electromotive force, several means were soon devised of preventing the formation of the film. Constant cells may be divided into two groups, according as their action is chemical (as in the bichromate cell, where the hydrogen is converted into water by an oxidizing agent placed in a porous pot round the carbon plate) or electrochemical (as in Daniell's cell, where a copper plate is surrounded by a solution of copper sulphate, and the hydrogen, instead of being liberated, replaces copper, which is deposited on the plate from the solution).

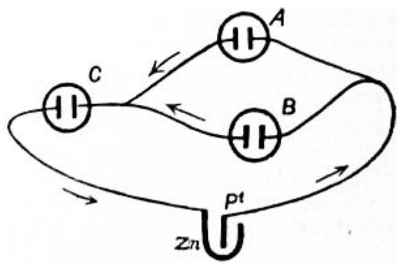

FIG. 1.

Faraday's Laws.—The first exact quantitative study of electrolytic phenomena was made about 1830 by Michael Faraday (*Experimental Researches*, 1833). When an electric current flows round a circuit, there is no accumulation of electricity anywhere in the circuit, hence the current strength is everywhere the same, and we may picture the current as analogous to the flow of an incompressible fluid. Acting on this view, Faraday set himself to examine the relation between the flow of electricity round the circuit and the amount of chemical decomposition. He passed the current driven by a voltaic battery ZnPt (fig. 1) through two branches containing the two electrolytic cells A and B. The reunited current was then led through another cell C, in which the strength of the current must be the sum of those in the arms A and B. Faraday found that the mass of substance liberated at the electrodes in the cell C was equal to the sum of the masses liberated in the cells A and B. He also found that, for the same current, the amount of chemical action was independent of the size of the electrodes and proportional to the time that the current flowed. Regarding the current as the passage of a certain amount of electricity per second, it will be seen that the results of all these experiments may be 218
summed up in the statement that the amount of chemical action is proportional to the quantity of electricity which passes through the cell.

Faraday's next step was to pass the same current through different electrolytes in series. He found that the amounts of the substances liberated in each cell were proportional to the chemical equivalent weights of those substances. Thus, if the current be passed through dilute sulphuric acid between hydrogen electrodes, and through a solution of copper sulphate, it will be found that the mass of hydrogen evolved in the first cell is to the mass of copper deposited in the second as 1 is to 31.8. Now this ratio is the same as that which gives the relative chemical equivalents of hydrogen and copper, for 1 gramme of hydrogen and 31.8 grammes of copper unite chemically with the same weight of any acid radicle such as chlorine or the sulphuric group, SO_4. Faraday examined also the electrolysis of certain fused salts such as lead chloride and silver chloride. Similar relations were found to hold and the amounts of chemical change to be the same for the same electric transfer as in the case of solutions.

We may sum up the chief results of Faraday's work in the statements known as Faraday's laws: The mass of substance liberated from an electrolyte by the passage of a current is proportional (1) to the total quantity of electricity which passes through the electrolyte, and (2) to the chemical equivalent weight of the substance liberated.

Since Faraday's time his laws have been confirmed by modern research, and in favourable cases have been shown to hold good with an accuracy of at least one part in a thousand. The principal object of this more recent research has been the determination of the quantitative amount of chemical change associated with the passage for a given time of a current of strength known in electromagnetic units. It is found that the most accurate and convenient apparatus to use is a platinum bowl filled with a solution of silver nitrate containing about fifteen parts of the salt to one hundred of water. Into the solution dips a silver plate wrapped in filter paper, and the current is passed from the silver plate as anode to the bowl as cathode. The bowl is weighed before and after the passage of the current, and the increase gives the mass of silver deposited. The mean result of the best determinations shows that when a current of one ampere is passed for one second, a mass of silver is deposited equal to 0.001118 gramme. So accurate and convenient is this determination that it is now used conversely as a practical definition of the ampere, which (defined theoretically in terms of magnetic force) is defined practically as the current which in one second deposits 1.118 milligramme of silver.

Taking the chemical equivalent weight of silver, as determined by chemical experiments, to be 107.92, the result described gives as the electrochemical equivalent of an ion of unit chemical equivalent the value 1.036×10^{-5}. If, as is now usual, we take the equivalent weight of oxygen as our standard and call it 16, the equivalent weight of hydrogen is 1.008, and its electrochemical equivalent is 1.044×10^{-5}. The electrochemical equivalent of any other substance, whether element or compound, may be found by multiplying its chemical equivalent by 1.036×10^{-5}. If, instead of the ampere, we take the C.G.S. electromagnetic unit of current, this number becomes 1.036×10^{-4}.

Chemical Nature of the Ions.—A study of the products of decomposition does not necessarily lead directly to a knowledge of the ions actually employed in carrying the current through the electrolyte. Since the electric forces are active throughout the whole solution, all the ions must come under its influence and therefore move, but their separation from the electrodes is determined by the electromotive force needed to liberate them. Thus, as long as every ion of the solution is present in the layer of liquid next the electrode, the one which responds to the least electromotive force will alone be set free. When the amount of this ion in the surface layer becomes too small to carry all the current across the junction, other ions must also be used, and either they or their secondary products will appear also at the electrode. In aqueous solutions, for instance, a few hydrogen (H) and hydroxyl (OH) ions derived from the water are always present, and will be liberated if the other ions require a higher decomposition voltage and the current be kept so small that hydrogen and hydroxyl ions can be formed fast enough to carry all the current across the junction between solution and electrode.

The issue is also obscured in another way. When the ions are set free at the electrodes, they may unite with the substance of the electrode or with some constituent of the solution to form secondary products. Thus the hydroxyl mentioned above decomposes into water and oxygen, and the chlorine produced by the electrolysis of a chloride may attack the metal of the anode. This leads us to examine more closely the part played by water in the electrolysis of aqueous solutions. Distilled water is a very bad conductor, though, even when great care is taken to remove all dissolved bodies, there is evidence to show that some part of the trace of conductivity remaining is due to the water itself. By careful redistillation F. Kohlrausch has prepared water of which the conductivity compared with that of mercury was only 0.40×10^{-11} at 18° C. Even here some little impurity was present, and the conductivity of chemically pure water was estimated by thermodynamic reasoning as 0.36×10^{-11} at 18° C. As we shall see later, the conductivity of very dilute salt solutions is proportional to the concentration, so that it is probable that, in most cases, practically all the current is carried by the salt. At the electrodes, however, the small quantity of hydrogen and hydroxyl ions from the water are liberated first in cases where the ions of the salt have a higher decomposition voltage. The water being present in excess, the hydrogen and hydroxyl are re-formed at once and therefore are set free continuously. If the current be so strong that new hydrogen and hydroxyl ions cannot be formed in time, other substances are liberated; in a solution of sulphuric acid a strong current will evolve sulphur dioxide, the more readily as the concentration of the solution is increased. Similar phenomena are seen in the case of a solution of hydrochloric acid. When the solution is weak, hydrogen and oxygen are evolved; but, as the concentration is increased, and the current raised, more and more chlorine is liberated.

An interesting example of secondary action is shown by the common technical process of electroplating with silver from a bath of potassium silver cyanide. Here the ions are potassium and the group $Ag(CN)_2$.[1] Each potassium ion as it reaches the cathode precipitates silver by reacting with the solution in accordance with the chemical equation

$$K + KAg(CN)_2 = 2KCN + Ag,$$

while the anion $Ag(CN)_2$ dissolves an atom of silver from the anode, and re-forms the complex cyanide $KAg(CN)_2$ by combining with the 2KCN produced in the reaction described in the equation. If the anode consist of platinum, cyanogen gas is evolved thereat from the anion $Ag(CN)_2$, and the platinum becomes covered with the insoluble silver cyanide, AgCN, which soon stops the current. The coating of silver obtained by this process is coherent and homogeneous, while that deposited from a solution of silver nitrate, as the result of the primary action of the current, is crystalline and easily detached.

In the electrolysis of a concentrated solution of sodium acetate, hydrogen is evolved at the cathode and a mixture of ethane and carbon dioxide at the anode. According to H. Jahn,[2] the processes at the anode can be represented by the equations

$$2CH_3{\cdot}COO + H_2O = 2CH_3{\cdot}COOH + O$$
$$2CH_3{\cdot}COOH + O = C_2H_6 + 2CO_2 + H_2O.$$

The hydrogen at the cathode is developed by the secondary action

$$2Na + 2H_2O = 2NaOH + H_2.$$

Many organic compounds can be prepared by taking advantage of secondary actions at the electrodes, such as reduction by the cathodic hydrogen, or oxidation at the anode (see Electrochemistry).

It is possible to distinguish between double salts and salts of compound acids. Thus J.W. Hittorf showed that when a current was passed through a solution of sodium platino-chloride, the platinum appeared at the anode. The salt must therefore be derived from an acid, chloroplatinic acid, H_2PtCl_6, and have the formula Na_2PtCl_6, the ions being Na and $PtCl_6''$, for if it were a double salt it would decompose as a mixture of sodium chloride and platinum chloride and both metals would go to the cathode.

Early Theories of Electrolysis.—The obvious phenomena to be explained by any theory of electrolysis are 219 the liberation of the products of chemical decomposition at the two electrodes while the intervening liquid is unaltered. To explain these facts, Theodor Grotthus (1785-1822) in 1806 put forward an hypothesis which supposed that the opposite chemical constituents of an electrolyte interchanged partners all along the line between the electrodes when a current passed. Thus, if the molecule of a substance in solution is represented by AB, Grotthus considered a chain of AB molecules to exist from one electrode to the other. Under the influence of an applied electric force, he imagined that the B part of the first molecule was liberated at the anode, and that the A part thus isolated united with the B part of the second molecule, which, in its turn, passed on its A to the B of the third molecule. In this manner, the B part of the last molecule of the chain was seized by the A of the last molecule but one, and the A part of the last molecule liberated at the surface of the cathode.

Chemical phenomena throw further light on this question. If two solutions containing the salts AB and CD be mixed, double decomposition is found to occur, the salts AD and CB being formed till a certain part of the first pair of substances is transformed into an equivalent amount of the second pair. The proportions between the four salts AB, CD, AD and CB, which exist finally in solution, are found to be the same whether we begin with the pair AB and CD or with the pair AD and CB. To explain this result, chemists suppose that both changes can occur simultaneously, and that equilibrium results when the rate at which AB and CD are transformed into AD and CB is the same as the rate at which the reverse change goes on. A freedom of interchange is thus indicated between the opposite parts of the molecules of salts in solution, and it follows reasonably that with the solution of a single salt, say sodium chloride, continual interchanges go on between the sodium and chlorine parts of the different molecules.

These views were applied to the theory of electrolysis by R.J.E. Clausius. He pointed out that it followed that the electric forces did not cause the interchanges between the opposite parts of the dissolved molecules but only controlled their direction. Interchanges must be supposed to go on whether a current passes or not, the function of the electric forces in electrolysis being merely to determine in what direction the parts of the molecules shall work their way through the liquid and to effect actual separation of these parts (or their secondary products) at the electrodes. This conclusion is supported also by the evidence supplied by the phenomena of electrolytic conduction (see Conduction, Electric, § II.). If we eliminate the reverse electromotive forces of polarization at the two electrodes, the conduction of electricity through electrolytes is found to conform to Ohm’s law; that is, once the polarization is overcome, the current is proportional to the electromotive force applied to the bulk of the liquid. Hence there can be no reverse forces of polarization inside the liquid itself, such forces being confined to the surface of the electrodes. No work is done in separating the parts of the molecules from each other. This result again indicates that the parts of the molecules are effectively separate from each other, the function of the electric forces being merely directive.

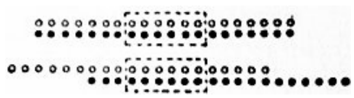

Fig. 2.

Migration of the Ions.—The opposite parts of an electrolyte, which work their way through the liquid under the action of the electric forces, were named by Faraday the ions—the travellers. The changes of concentration which occur in the solution near the two electrodes were referred by W. Hittorf (1853) to the unequal speeds with which he supposed the two opposite ions to travel. It is clear that, when two opposite streams of ions move past each other, equivalent quantities are liberated at the two ends of the system. If the ions move at equal rates, the salt which is decomposed to supply the ions liberated must be taken equally from the neighbourhood of the two electrodes. But if one ion, say the anion, travels faster through the liquid than the other, the end of the solution from which it comes will be more exhausted of salt than the end towards which it goes. If we assume that no other cause is at work, it is easy to prove that, with non-dissolvable electrodes, the ratio of salt lost at the anode to the salt lost at the cathode must be equal to the ratio of the velocity of the cation to the velocity of the anion. This result may be illustrated by fig. 2. The black circles represent one ion and the white circles the other. If the black ions move twice as fast as the white ones, the state of things after the passage of a current will be represented by the lower part of the figure. Here the middle part of the solution is unaltered and the number of ions liberated is the same at either end, but the amount of salt left at one end is less than that at the other. On the right, towards which the faster ion travels, five molecules of salt are left, being a loss of two from the original seven. On the left, towards which the slower ion moves, only three molecules remain—a loss of four. Thus, the ratio of the losses at the two ends is two to one—the same as the ratio of the assumed ionic velocities. It should be noted, however, that another cause would be competent to explain the unequal dilution of the two solutions. If either ion carried with it some of the unaltered salt or some of the solvent, concentration or dilution of the liquid would be produced where the ion was liberated. There is reason to believe that in certain cases such complex ions do exist, and interfere with the results of the differing ionic velocities.

Hittorf and many other observers have made experiments to determine the unequal dilution of a solution round the two electrodes when a current passes. Various forms of apparatus have been used, the principle of them all being to secure efficient separation of the two volumes of solution in which the changes occur. In some cases porous diaphragms have been employed; but such diaphragms introduce a new complication, for the liquid as a whole is pushed through them by the action of the current, the phenomenon being known as electric endosmose. Hence experiments without separating diaphragms are to be preferred, and the apparatus may be considered effective when a considerable bulk of intervening solution is left unaltered in composition. It is usual to express the results in terms of what is called the migration constant of the anion, that is, the ratio of the amount of salt lost by the anode vessel to the whole amount lost by both vessels. Thus the statement that the migration constant or transport number for a decinormal solution of copper sulphate is 0.632 implies that of every gramme of copper sulphate lost by a solution containing originally one-tenth of a gramme equivalent per litre when a current is passed through it between platinum electrodes, 0.632 gramme is taken from the cathode vessel and 0.368 gramme from the anode vessel. For certain concentrated solutions the transport number is found to be greater than unity; thus for a normal solution of cadmium iodide its value is 1.12. On the theory that the phenomena are wholly due to unequal ionic velocities this result would mean that the cation like the anion moved against the conventional direction of the current. That a body carrying a positive electric charge should move against the direction of the electric intensity is contrary to all our notions of electric forces, and we are compelled to seek some other explanation. An alternative hypothesis is given by the idea of complex ions. If some of the anions, instead of being simple iodine ions represented chemically by the symbol I, are complex structures formed by the union of iodine with unaltered cadmium iodide—structures represented by some such chemical formula as $I(CdI_2)$, the concentration of the solution round the anode would be increased by the passage of an electric current, and the phenomena observed would be explained. It is found that, in such cases as this, where it seems necessary to imagine the existence of complex ions, the transport number changes rapidly as the concentration of the original solution is changed. Thus, diminishing the concentration of the cadmium iodine solution from normal to one-twentieth normal changes the transport number from 1.12 to 0.64. Hence it is probable that in cases where the transport number keeps constant with changing concentration the hypothesis of complex ions is 220
unnecessary, and we may suppose that the transport number is a true migration constant from which the relative velocities of the two ions may be calculated in the matter suggested by Hittorf and illustrated in fig. 2. This conclusion is confirmed by the results of the direct visual determination of ionic velocities (see Conduction, Electric, § II.), which, in cases where the transport number remains constant, agree with the values calculated from those numbers. Many solutions in which the transport numbers vary at high concentration often become simple at greater dilution. For instance, to take the two solutions to which we have already referred, we have—

Concentration	2.0	1.5	1.0	0.5	0.2	0.1	0.05	0.02	0.01 normal
Copper sulphate transport numbers	0.72	0.714	0.696	0.668	0.643	0.632	0.626	0.62	. .
Cadmium iodide ” ”	1.22	1.18	1.12	1.00	0.83	0.71	0.64	0.59	0.56

It is probable that in both these solutions complex ions exist at fairly high concentrations, but gradually gets less in number and finally disappear as the dilution is increased. In such salts as potassium chloride the ions seem to be simple throughout a wide range of concentration since the transport numbers for the same series of concentrations as those used above run—

Potassium chloride—
0.515, 0.515, 0.514, 0.513, 0.509, 0.508, 0.507, 0.507, 0.506.

The next important step in the theory of the subject was made by F. Kohlrausch in 1879. Kohlrausch formulated a theory of electrolytic conduction based on the idea that, under the action of the electric forces, the oppositely charged ions moved in opposite directions through the liquid, carrying their charges with them. If we eliminate the polarization at the electrodes, it can be shown that an electrolyte possesses a definite electric resistance and therefore a definite conductivity. The conductivity gives us the amount of electricity conveyed per second under a definite electromotive force. On the view of the process of conduction described above, the amount of electricity conveyed per second is measured by the product of the number of ions, known from the concentration of the solution, the charge carried by each of them, and the velocity with which, on the average, they move through the liquid. The concentration is known, and the conductivity can be measured experimentally; thus the average velocity with which the ions move past each other under the existent electromotive force can be estimated. The velocity with which the ions move past each other is equal to the sum of their individual velocities, which can therefore be calculated. Now Hittorf’s transport number, in the case of simple salts in moderately dilute solution, gives us the ratio between the two ionic velocities. Hence the absolute velocities of the two ions can be determined, and we can calculate the actual speed with which a certain ion moves through a given liquid under the action of a given potential gradient or electromotive force. The details of the calculation are given in the article Conduction, Electric, § II., where also will be found an account of the methods which have been used to measure the velocities of many ions by direct visual observation. The results go to show that, where the existence of complex ions is not indicated by varying transport numbers, the observed velocities agree with those calculated on Kohlrausch’s theory.

Dissociation Theory.—The verification of Kohlrausch’s theory of ionic velocity verifies also the view of electrolysis which regards the electric current as due to streams of ions moving in opposite directions through the liquid and carrying their opposite electric charges with them. There remains the question how the necessary migratory freedom of the ions is secured. As we have seen, Grotthus imagined that it was the electric forces which sheared the ions past each other and loosened the chemical bonds holding the opposite parts of each dissolved molecule together. Clausius extended to electrolysis the chemical ideas which looked on the opposite parts of the molecule as always changing partners independently of any electric force, and regarded the function of the current as merely directive. Still, the necessary freedom was supposed to be secured by interchanges of ions between molecules at the instants of molecular collision only; during the rest of the life of the ions they were regarded as linked to each other to form electrically neutral molecules.

In 1887 Svante Arrhenius, professor of physics at Stockholm, put forward a new theory which supposed that the freedom of the opposite ions from each other was not a mere momentary freedom at the instants of molecular collision, but a more or less permanent freedom, the ions moving independently of each other through the liquid. The evidence which led Arrhenius to this conclusion was based on van 't Hoff's work on the osmotic pressure of solutions (see Solution). If a solution, let us say of sugar, be confined in a closed vessel through the walls of which the solvent can pass but the solution cannot, the solvent will enter till a certain equilibrium pressure is reached. This equilibrium pressure is called the osmotic pressure of the solution, and thermodynamic theory shows that, in an ideal case of perfect separation between solvent and solute, it should have the same value as the pressure which a number of molecules equal to the number of solute molecules in the solution would exert if they could exist as a gas in a space equal to the volume of the solution, provided that the space was large enough (*i.e.* the solution dilute enough) for the intermolecular forces between the dissolved particles to be inappreciable. Van 't Hoff pointed out that measurements of osmotic pressure confirmed this value in the case of dilute solutions of cane sugar.

Thermodynamic theory also indicates a connexion between the osmotic pressure of a solution and the depression of its freezing point and its vapour pressure compared with those of the pure solvent. The freezing points and vapour pressures of solutions of sugar are also in conformity with the theoretical numbers. But when we pass to solutions of mineral salts and acids—to solutions of electrolytes in fact—we find that the observed values of the osmotic pressures and of the allied phenomena are greater than the normal values. Arrhenius pointed out that these exceptions would be brought into line if the ions of electrolytes were imagined to be separate entities each capable of producing its own pressure effects just as would an ordinary dissolved molecule.

Two relations are suggested by Arrhenius' theory. (1) In very dilute solutions of simple substances, where only one kind of dissociation is possible and the dissociation of the ions is complete, the number of pressure-producing particles necessary to produce the observed osmotic effects should be equal to the number of ions given by a molecule of the salt as shown by its electrical properties. Thus the osmotic pressure, or the depression of the freezing point of a solution of potassium chloride should, at extreme dilution, be twice the normal value, but of a solution of sulphuric acid three times that value, since the potassium salt contains two ions and the acid three. (2) As the concentration of the solutions increases, the ionization as measured electrically and the dissociation as measured osmotically might decrease more or less together, though, since the thermodynamic theory only holds when the solution is so dilute that the dissolved particles are beyond each other's sphere of action, there is much doubt whether this second relation is valid through any appreciable range of concentration.

At present, measurements of freezing point are more convenient and accurate than those of osmotic pressure, and we may test the validity of Arrhenius' relations by their means. The theoretical value for the depression of the freezing point of a dilute solution per gramme-equivalent of solute per litre is 1.857° C. Completely ionized solutions of salts with two ions should give double this number or 3.714°, while electrolytes with three ions should have a value of 5.57°.

The following results are given by H.B. Loomis for the concentration of 0.01 gramme-molecule of salt to one thousand grammes of water. The salts tabulated are those of which the equivalent conductivity reaches a 221
limiting value indicating that complete ionization is reached as dilution is increased. With such salts alone is a valid comparison possible.

Molecular Depressions of the Freezing Point.

Electrolytes with two Ions.

Potassium chloride	3.60	Nitric acid	3.73
Sodium chloride	3.67	Potassium nitrate	3.46
Potassium hydrate	3.71	Sodium nitrate	3.55
Hydrochloric acid	3.61	Ammonium nitrate	3.58

Electrolytes with three Ions.

Sulphuric acid	4.49	Calcium chloride	5.04
Sodium sulphate	5.09	Magnesium chloride	5.08

At the concentration used by Loomis the electrical conductivity indicates that the ionization is not complete, particularly in the case of the salts with divalent ions in the second list. Allowing for incomplete ionization the general concordance of these numbers with the theoretical ones is very striking.

The measurements of freezing points of solutions at the extreme dilution necessary to secure complete ionization is a matter of great difficulty, and has been overcome only in a research initiated by E.H. Griffiths.[3] Results have been obtained for solutions of sugar, where the experimental number is 1.858, and for potassium chloride, which gives a depression of 3.720. These numbers agree with those indicated by theory, viz. 1.857 and 3.714, with astonishing exactitude. We may take Arrhenius' first relation as established for the case of potassium chloride.

The second relation, as we have seen, is not a strict consequence of theory, and experiments to examine it must be treated as an investigation of the limits within which solutions are dilute within the thermodynamic sense of the word, rather than as a test of the soundness of the theory. It is found that divergence has begun before the concentration has become great enough to enable freezing points to be measured with any ordinary apparatus. The freezing point curve usually lies below the electrical one, but approaches it as dilution is increased.[4]

Returning once more to the consideration of the first relation, which deals with the comparison between the number of ions and the number of pressure-producing particles in dilute solution, one caution is necessary. In simple substances like potassium chloride it seems evident that one kind of dissociation only is possible. The electrical phenomena show that there are two ions to the molecule, and that these ions are electrically charged. Corresponding with this result we find that the freezing point of dilute solutions indicates that two pressure-producing particles per molecule are present. But the converse relation does not necessarily follow. It would be possible for a body in solution to be dissociated into non-electrical parts, which would give osmotic pressure effects twice or three times the normal value, but, being uncharged, would not act as ions and impart electrical conductivity to the solution. L. Kahlenberg (*Jour. Phys. Chem.*, 1901, v. 344, 1902, vi. 43) has found that solutions of diphenylamine in methyl cyanide possess an excess of pressure-producing particles and yet are non-conductors of electricity. It is possible that in complicated organic substances we might have two kinds of dissociation, electrical and non-electrical, occurring simultaneously, while the possibility of the association of molecules accompanied by the electrical dissociation of some of them into new parts should not be overlooked. It should be pointed out that no measurements on osmotic pressures or freezing points can do more than tell us that an excess of particles is present; such experiments can throw no light on the question whether or not those particles are electrically charged. That question can only be answered by examining whether or not the particles move in an electric field.

The dissociation theory was originally suggested by the osmotic pressure relations. But not only has it explained satisfactorily the electrical properties of solutions, but it seems to be the only known hypothesis which is consistent with the experimental relation between the concentration of a solution and its electrical conductivity (see Conduction, Electric, § II., "Nature of Electrolytes"). It is probable that the electrical effects constitute the strongest arguments in favour of the theory. It is necessary to point out that the dissociated ions of such a body as potassium chloride are not in the same condition as potassium and chlorine in the free state. The ions are associated with very large electric charges, and, whatever their exact relations with those charges may be, it is certain that the energy of a system in such a state must be different from its energy when unelectrified. It is not unlikely, therefore, that even a compound as stable in the solid form as potassium chloride should be thus dissociated when dissolved. Again, water, the best electrolytic solvent known, is also the body of the highest specific inductive capacity (dielectric constant), and this property, to whatever cause it may be due, will reduce the forces between electric charges in the neighbourhood, and may therefore enable two ions to separate.

This view of the nature of electrolytic solutions at once explains many well-known phenomena. Other physical properties of these solutions, such as density, colour, optical rotatory power, &c., like the conductivities, are *additive*, *i.e.* can be calculated by adding together the corresponding properties of the parts. This again suggests that these parts are independent of each other. For instance, the colour of a salt solution is the colour obtained by the superposition of the colours of the ions and the colour of any undissociated salt that may be present. All copper salts in dilute solution are blue, which is therefore the colour of the copper ion. Solid copper chloride is brown or yellow, so that its concentrated solution, which contains both ions and undissociated molecules, is green, but changes to blue as water is added and the ionization becomes complete. A series of equivalent solutions all containing the same coloured ion have absorption spectra which, when photographed, show identical absorption bands of equal intensity.[5] The colour changes shown by many substances which are used as indicators (*q.v.*) of acids or alkalis can be explained in a similar way. Thus para-nitrophenol has colourless molecules, but an intensely yellow negative ion. In neutral, and still more in acid solutions, the dissociation of the indicator is practically nothing, and the liquid is colourless. If an alkali is added, however, a highly dissociated salt of para-nitrophenol is formed, and the yellow colour is at once evident. In other cases, such as that of litmus, both the ion and the undissociated molecule are coloured, but in different ways.

Electrolytes possess the power of coagulating solutions of colloids such as albumen and arsenious sulphide. The mean values of the relative coagulative powers of sulphates of mono-, di-, and tri-valent metals have been shown experimentally to be approximately in the ratios 1 : 35 : 1023. The dissociation theory refers this to the action of electric charges carried by the free ions. If a certain minimum charge must be collected in order to start coagulation, it will need the conjunction of 6n monovalent, or 3n divalent, to equal the effect of 2n tri-valent ions. The ratios of the coagulative powers can thus be calculated to be $1 : x : x^2$, and putting $x = 32$ we get 1 : 32 : 1024, a satisfactory agreement with the numbers observed.[6]

The question of the application of the dissociation theory to the case of fused salts remains. While it seems clear that the conduction in this case is carried on by ions similar to those of solutions, since Faraday's laws apply equally to both, it does not follow necessarily that semi-permanent dissociation is the only way to explain the phenomena. The evidence in favour of dissociation in the case of solutions does not apply to fused salts, and it is possible that, in their case, a series of molecular interchanges, somewhat like Grotthus's chain, may represent the mechanism of conduction.

An interesting relation appears when the electrolytic conductivity of solutions is compared with their chemical activity. The readiness and speed with which electrolytes react are in sharp contrast with the difficulty experienced in the case of non-electrolytes. Moreover, a study of the chemical relations of electrolytes indicates that it is always the electrolytic ions that are concerned in their reactions. The tests for a salt, potassium nitrate, for example, are the tests not for KNO_3, but for its ions K and NO_3, and in cases of double decomposition it is always these ions that are exchanged for those of other substances. If an element be present in a compound otherwise than as an ion, it is not interchangeable, and cannot be recognized by the usual tests. Thus neither a chlorate, which contains the ion ClO_3, nor monochloracetic acid, shows the reactions of chlorine, though it is, of course, present in both substances; again, the sulphates do not answer to the usual tests which indicate the presence of sulphur as sulphide. The chemical activity of a substance is a quantity which may be measured by different methods. For some substances it has been shown to be independent of the particular reaction used. It is then possible to assign to each body a specific coefficient of 222 affinity. Arrhenius has pointed out that the coefficient of affinity of an acid is proportional to its electrolytic ionization.

The affinities of acids have been compared in several ways. W. Ostwald (*Lehrbuch der allg. Chemie*, vol. ii., Leipzig, 1893) investigated the relative affinities of acids for potash, soda and ammonia, and proved them to be independent of the base used. The method employed was to measure the changes in volume caused by the action. His results are given in column I. of the following table, the affinity of hydrochloric acid being taken as one hundred. Another method is to allow an acid to act on an insoluble salt, and to measure the quantity which goes into solution. Determinations have been made with calcium oxalate, $CaC_2O_4 + H_2O$, which is easily decomposed by acids, oxalic acid and a soluble calcium salt being formed. The affinities of acids relative to that of oxalic acid are thus found, so that the acids can be compared among themselves (column II.). If an aqueous solution of methyl acetate be allowed to stand, a slow decomposition goes on. This is much quickened by the presence of a little dilute acid, though the acid itself remains unchanged. It is found that the influence of different acids on this action is proportional to their specific coefficients of affinity. The results of this method are given in column III. Finally, in column IV. the electrical conductivities of normal solutions of the acids have been tabulated. A better basis of comparison would be the ratio of the actual to the limiting conductivity, but since the conductivity of acids is chiefly due to the mobility of the hydrogen ions, its limiting value is nearly the same for all, and the general result of the comparison would be unchanged.

Acid.	I.	II.	III.	IV.
Hydrochloric	100	100	100	100
Nitric	102	110	92	99.6
Sulphuric	68	67	74	65.1
Formic	4.0	2.5	1.3	1.7
Acetic	1.2	1.0	0.3	0.4
Propionic	1.1	· ·	0.3	0.3
Monochloracetic	7.2	5.1	4.3	4.9
Dichloracetic	34	18	23.0	25.3
Trichloracetic	82	63	68.2	62.3
Malic	3.0	5.0	1.2	1.3
Tartaric	5.3	6.3	2.3	2.3
Succinic	0.1	0.2	0.5	0.6

It must be remembered that, the solutions not being of quite the same strength, these numbers are not strictly comparable, and that the experimental difficulties involved in the chemical measurements are considerable. Nevertheless, the remarkable general agreement of the numbers in the four columns is quite enough to show the intimate connexion between chemical activity and electrical conductivity. We may take it, then, that only that portion of these bodies is chemically active which is electrolytically active—that ionization is necessary for such chemical activity as we are dealing with here, just as it is necessary for electrolytic conductivity.

The ordinary laws of chemical equilibrium have been applied to the case of the dissociation of a substance into its ions. Let x be the number of molecules which dissociate per second when the number of undissociated molecules in unit volume is unity, then in a dilute solution where the molecules do not interfere with each other, xp is the number when the concentration is p. Recombination can only occur when two ions meet, and since the frequency with which this will happen is, in dilute solution, proportional to the square of the ionic concentration, we shall get for the number of molecules re-formed in one second yq^2 where q is the number of dissociated molecules in one cubic centimetre. When there is equilibrium, $xp = yq^2$. If μ be the molecular conductivity, and μ_∞ its value at infinite dilution, the fractional number of molecules dissociated is μ / μ_∞, which we may write as α. The number of undissociated molecules is then 1 − α, so that if V be the volume of the solution containing 1 gramme-molecule of the dissolved substance, we get

$$q = \alpha / V \text{ and } p = (1 - \alpha) / V,$$

hence

$$x (1 - \alpha) V = ya^2 / V^2,$$

and

$$\frac{\alpha^2}{V (1 - \alpha)} = \frac{x}{y} = \text{constant} = k.$$

This constant k gives a numerical value for the chemical affinity, and the equation should represent the effect of dilution on the molecular conductivity of binary electrolytes.

In the case of substances like ammonia and acetic acid, where the dissociation is very small, 1 − α is nearly equal to unity, and only varies slowly with dilution. The equation then becomes $\alpha^2/V = k$, or $\alpha = \sqrt{(Vk)}$, so that the molecular conductivity is proportional to the square root of the dilution. Ostwald has confirmed the equation by observation on an enormous number of weak acids (*Zeits. physikal. Chemie*, 1888, ii. p. 278; 1889, iii. pp. 170, 241, 369). Thus in the case of cyanacetic acid, while the volume V changed by doubling from 16 to 1024 litres, the values of k were 0.00 (376, 373, 374, 361, 362, 361, 368). The mean values of k for other common acids were—formic, 0.0000214; acetic, 0.0000180; monochloracetic, 0.00155; dichloracetic, 0.051; trichloracetic, 1.21; propionic, 0.0000134. From these numbers we can, by help of the equation, calculate the conductivity of the acids for any dilution. The value of k, however, does not keep constant so satisfactorily in the case of highly dissociated substances, and empirical formulae have been constructed to represent the effect of dilution on them. Thus the values of the expressions $\alpha^2 / (1 - \alpha\sqrt{V})$ (Rudolphi, *Zeits. physikal. Chemie*, 1895, vol. xvii. p. 385) and $\alpha^3 / (1 - \alpha)^2 V$ (van 't Hoff, ibid., 1895, vol. xviii. p. 300) are found to keep constant as V changes. Van 't Hoff's formula is equivalent to taking the frequency of dissociation as proportional to the square of the concentration of the molecules, and the frequency of recombination as proportional to the cube of the concentration of the ions. An explanation of the failure of the usual dilution law in these cases may be given if we remember that, while the electric forces between bodies like undissociated molecules, each associated with equal and opposite charges, will vary inversely as the fourth power of the distance, the forces between dissociated ions, each carrying one charge only, will be inversely proportional to the square of the distance. The forces between the ions of a strongly dissociated solution will thus be considerable at a dilution which makes forces between undissociated molecules quite insensible, and at the concentrations necessary to test Ostwald's formula an electrolyte will be far from dilute in the thermodynamic sense of the term, which implies no appreciable intermolecular or interionic forces.

When the solutions of two substances are mixed, similar considerations to those given above enable us to calculate the resultant changes in dissociation. (See Arrhenius, *loc. cit.*) The simplest and most important case is that of two electrolytes having one ion in common, such as two acids. It is evident that the undissociated part of each acid must eventually be in equilibrium with the free hydrogen ions, and, if the concentrations are not such as to secure this condition, readjustment must occur. In order that there should be no change in the states of dissociation on mixing, it is necessary, therefore, that the concentration of the hydrogen ions should be the same in each separate solution. Such solutions were called by Arrhenius "isohydric." The two solutions, then, will so act on each other when mixed that they become isohydric. Let us suppose that we have one very active acid like hydrochloric, in which dissociation is nearly complete, another like acetic, in which it is very small. In order that the solutions of these should be isohydric and the concentrations of the hydrogen ions the same, we must have a very large quantity of the feebly dissociated acetic acid, and a very small quantity of the strongly dissociated hydrochloric, and in such proportions alone will equilibrium be possible. This explains the action of a strong acid on the salt of a weak acid. Let us allow dilute sodium acetate to react with dilute hydrochloric acid. Some acetic acid is formed, and this process will go on till the solutions of the two acids are isohydric: that is, till the dissociated hydrogen ions are in equilibrium with both. In order that this should hold, we have seen that a considerable quantity of acetic acid must be present, so that a corresponding amount of the salt will be decomposed, the quantity being greater the less the acid is dissociated. This "replacement" of a "weak" acid by a "strong" one is a matter of common observation in the chemical laboratory. Similar investigations applied to the general case of chemical equilibrium lead to an expression of exactly the same form as that given by C.M. Guldberg and P. Waage, which is universally accepted as an accurate representation of the facts.

The temperature coefficient of conductivity has approximately the same value for most aqueous salt solutions. It decreases both as the temperature is raised and as the concentration is increased, ranging from about 3.5% per degree for extremely dilute solutions (*i.e.* practically pure water) at 0° to about 1.5 for 223
concentrated solutions at 18°. For acids its value is usually rather less than for salts at equivalent concentrations. The influence of temperature on the conductivity of solutions depends on (1) the ionization, and (2) the frictional resistance of the liquid to the passage of the ions, the reciprocal of which is called the ionic fluidity. At extreme dilution, when the ionization is complete, a variation in temperature cannot change its amount. The rise of conductivity with temperature, therefore, shows that the fluidity becomes greater when the solution is heated. As the concentration is increased and un-ionized molecules are formed, a change in temperature begins to affect the ionization as well as the fluidity. But the temperature coefficient of conductivity is now generally less than before; thus the effect of temperature on ionization must be of opposite sign to its effect on fluidity. The ionization of a solution, then, is usually diminished by raising the temperature, the rise in conductivity being due to the greater increase in fluidity. Nevertheless, in certain cases, the temperature coefficient of conductivity becomes negative at high temperatures, a solution of phosphoric acid, for example, reaching a maximum conductivity at 75° C.

The dissociation theory gives an immediate explanation of the fact that, in general, no heat-change occurs when two neutral salt solutions are mixed. Since the salts, both before and after mixture, exist mainly as dissociated ions, it is obvious that large thermal effects can only appear when the state of dissociation of the products is very different from that of the reagents. Let us consider the case of the neutralization of a base by an acid in the light of the dissociation theory. In dilute solution such substances as hydrochloric acid and potash are almost completely dissociated, so that, instead of representing the reaction as

$$HCl + KOH = KCl + H_2O,$$

we must write

$$\overset{+}{H} + \overset{-}{Cl} + \overset{+}{K} + \overset{-}{OH} = \overset{+}{K} + \overset{-}{Cl} + H_2O.$$

The ions K and Cl suffer no change, but the hydrogen of the acid and the hydroxyl (OH) of the potash unite to form water, which is only very slightly dissociated. The heat liberated, then, is almost exclusively that produced by the formation of water from its ions. An exactly similar process occurs when any strongly dissociated acid acts on any strongly dissociated base, so that in all such cases the heat evolution should be approximately the same. This is fully borne out by the experiments of Julius Thomsen, who found that the heat of neutralization of one gramme-molecule of a strong base by an equivalent quantity of a strong acid was nearly constant, and equal to 13,700 or 13,800 calories. In the case of weaker acids, the dissociation of which is less complete, divergences from this constant value will occur, for some of the molecules have to be separated into their ions. For instance, sulphuric acid, which in the fairly strong solutions used by Thomsen is only about half dissociated, gives a higher value for the heat of neutralization, so that heat must be evolved when it is ionized. The heat of formation of a substance from its ions is, of course, very different from that evolved when it is formed from its elements in the usual way, since the energy associated with an ion is different from that possessed by the atoms of the element in their normal state. We can calculate the heat of formation from its ions for any substance dissolved in a given liquid, from a knowledge of the temperature coefficient of ionization, by means of an application of the well-known thermodynamical process, which also gives the latent heat of evaporation of a liquid when the temperature coefficient of its vapour pressure is known. The heats of formation thus obtained may be either positive or negative, and by using them to supplement the heat of formation of water, Arrhenius calculated the total heats of neutralization of soda by different acids, some of them only slightly dissociated, and found values agreeing well with observation (*Zeits. physikal. Chemie*, 1889, 4, p. 96; and 1892, 9, p. 339).

Voltaic Cells.—When two metallic conductors are placed in an electrolyte, a current will flow through a wire connecting them provided that a difference of any kind exists between the two conductors in the nature either of the metals or of the portions of the electrolyte which surround them. A current can be obtained by the combination of two metals in the same electrolyte, of two metals in different electrolytes, of the same metal in different electrolytes, or of the same metal in solutions of the same electrolyte at different concentrations. In accordance with the principles of energetics (*q.v.*), any change which involves a decrease in the total available energy of the system will tend to occur, and thus the necessary and sufficient condition for the production of electromotive force is that the available energy of the system should decrease when the current flows.

In order that the current should be maintained, and the electromotive force of the cell remain constant during action, it is necessary to ensure that the changes in the cell, chemical or other, which produce the current, should neither destroy the difference between the electrodes, nor coat either electrode with a non-conducting layer through which the current cannot pass. As an example of a fairly constant cell we may take that of Daniell, which consists of the electrical arrangement—zinc | zinc sulphate solution | copper sulphate solution | copper,—the two solutions being usually separated by a pot of porous earthenware. When the zinc and copper plates are connected through a wire, a current flows, the conventionally positive electricity passing from copper to zinc in the wire and from zinc to copper in the cell. Zinc dissolves at the anode, an equal amount of zinc replaces an equivalent amount of copper on the other side of the porous partition, and the same amount of copper is deposited on the cathode. This process involves a decrease in the available energy of the system, for the dissolution of zinc gives out more energy than the separation of copper absorbs. But the internal rearrangements which accompany the production of a current do not cause any change in the original nature of the electrodes, fresh zinc being exposed at the anode, and copper being deposited on copper at the cathode. Thus as long as a moderate current flows, the only variation in the cell is the appearance of zinc sulphate in the liquid on the copper side of the porous wall. In spite of this appearance, however, while the supply of copper is maintained, copper, being more easily separated from the solution than zinc, is deposited alone at the cathode, and the cell remains constant.

It is necessary to observe that the condition for change in a system is that the total available energy of the whole system should be decreased by the change. We must consider what change is allowed by the mechanism of the system, and deal with the sum of all the alterations in energy. Thus in the Daniell cell the dissolution of copper as well as of zinc would increase the loss in available energy. But when zinc dissolves, the zinc ions carry their electric charges with them, and the liquid tends to become positively electrified. The electric forces then soon stop further action unless an equivalent quantity of positive ions are removed from the solution. Hence zinc can only dissolve when some more easily separable substance is present in solution to be removed pari passu with the dissolution of zinc. The mechanism of such systems is well illustrated by an experiment devised by W. Ostwald. Plates of platinum and pure or amalgamated zinc are separated by a porous pot, and each surrounded by some of the same solution of a salt of a metal more oxidizable than zinc, such as potassium. When the plates are connected together by means of a wire, no current flows, and no appreciable amount of zinc dissolves, for the dissolution of zinc would involve the separation of potassium and a gain in available energy. If sulphuric acid be added to the vessel containing the zinc, these conditions are unaltered and still no zinc is dissolved. But, on the other hand, if a few drops of acid be placed in the vessel with the platinum, bubbles of hydrogen appear, and a current flows, zinc dissolving at the anode, and hydrogen being liberated at the cathode. In order that positively electrified ions may enter a solution, an equivalent amount of other positive ions must be removed or negative ions be added, and, for the process to occur spontaneously, the possible action at the two electrodes must involve a decrease in the total available energy of the system.

Considered thermodynamically, voltaic cells must be divided into reversible and non-reversible systems. If 224
the slow processes of diffusion be ignored, the Daniell cell already described may be taken as a type of a reversible cell. Let an electromotive force exactly equal to that of the cell be applied to it in the reverse direction. When the applied electromotive force is diminished by an infinitesimal amount, the cell produces a current in the usual direction, and the ordinary chemical changes occur. If the external electromotive force exceed that of the cell by ever so little, a current flows in the opposite direction, and all the former chemical changes are reversed, copper dissolving from the copper plate, while zinc is deposited on the zinc plate. The cell, together with this balancing electromotive force, is thus a reversible system in true equilibrium, and the thermodynamical reasoning applicable to such systems can be used to examine its properties.

Now a well-known relation connects the available energy of a reversible system with the corresponding change in its total internal energy.

The available energy A is the amount of external work obtainable by an infinitesimal, reversible change in the system which occurs at a constant temperature T. If I be the change in the internal energy, the relation referred to gives us the equation

$$A = I + T\,(dA/dT),$$

where dA/dT denotes the rate of change of the available energy of the system per degree change in temperature. During a small electric transfer through the cell, the external work done is Ee, where E is the electromotive force. If the chemical changes which occur in the cell were allowed to take place in a closed vessel without the performance of electrical or other work, the change in energy would be measured by the heat evolved. Since the final state of the system would be the same as in the actual processes of the cell, the same amount of heat must give a measure of the change in internal energy when the cell is in action. Thus, if L denote the heat corresponding with the chemical changes associated with unit electric transfer, Le will be the heat corresponding with an electric transfer e, and will also be equal to the change in internal energy of the cell. Hence we get the equation

$$Ee = Le + Te\,(dE/dT) \text{ or } E = L + T\,(dE/dT),$$

as a particular case of the general thermodynamic equation of available energy. This equation was obtained in different ways by J. Willard Gibbs and H. von Helmholtz.

It will be noticed that when dE/dT is zero, that is, when the electromotive force of the cell does not change with temperature, the electromotive force is measured by the heat of reaction per unit of electrochemical change. The earliest formulation of the subject, due to Lord Kelvin, assumed that this relation was true in all cases, and, calculated in this way, the electromotive force of Daniell's cell, which happens to possess a very small temperature coefficient, was found to agree with observation.

When one gramme of zinc is dissolved in dilute sulphuric acid, 1670 thermal units or calories are evolved. Hence for the electrochemical unit of zinc or 0.003388 gramme, the thermal evolution is 5.66 calories. Similarly, the heat which accompanies the dissolution of one electrochemical unit of copper is 3.00 calories. Thus, the thermal equivalent of the unit of resultant electrochemical change in Daniell's cell is 5.66 − 3.00 = 2.66 calories. The dynamical equivalent of the calorie is 4.18×10^7 ergs or C.G.S. units of work, and therefore the electromotive force of the cell should be 1.112×10^8 C.G.S. units or 1.112 volts—a close agreement with the experimental result of about 1.08 volts. For cells in which the electromotive force varies with temperature, the full equation given by Gibbs and Helmholtz has also been confirmed experimentally.

As stated above, an electromotive force is set up whenever there is a difference of any kind at two electrodes immersed in electrolytes. In ordinary cells the difference is secured by using two dissimilar metals, but an electromotive force exists if two plates of the same metal are placed in solutions of different substances, or of the same substance at different concentrations. In the latter case, the tendency of the metal to dissolve in the more dilute solution is greater than its tendency to dissolve in the more concentrated solution, and thus there is a decrease in available energy when metal dissolves in the dilute solution and separates in equivalent quantity from the concentrated solution. An electromotive force is therefore set up in this direction, and, if we can calculate the change in available energy due to the processes of the cell, we can foretell the value of the electromotive force. Now the effective change produced by the action of the current is the concentration of the more dilute solution by the dissolution of metal in it, and the dilution of the originally stronger solution by the separation of metal from it. We may imagine these changes reversed in two ways. We may evaporate some of the solvent from the solution which has become weaker and thus reconcentrate it, condensing the vapour on the solution which had become stronger. By this reasoning Helmholtz showed how to obtain an expression for the work done. On the other hand, we may imagine the processes due to the electrical transfer to be reversed by an osmotic operation. Solvent may be supposed to be squeezed out from the solution which has become more dilute through a semi-permeable wall, and through another such wall allowed to mix with the solution which in the electrical operation had become more concentrated. Again, we may calculate the osmotic work done, and, if the whole cycle of operations be supposed to occur at the same temperature, the osmotic work must be equal and opposite to the electrical work of the first operation.

The result of the investigation shows that the electrical work Ee is given by the equation

$$Ee = \int_{p1}^{p2} vdp,$$

where v is the volume of the solution used and p its osmotic pressure. When the solutions may be taken as effectively dilute, so that the gas laws apply to the osmotic pressure, this relation reduces to

$$E = \frac{nrRT}{ey} \log_\varepsilon \frac{c_1}{c_2}$$

where n is the number of ions given by one molecule of the salt, r the transport ratio of the anion, R the gas constant, T the absolute temperature, y the total valency of the anions obtained from one molecule, and c_1 and c_2 the concentrations of the two solutions.

If we take as an example a concentration cell in which silver plates are placed in solutions of silver nitrate, one of which is ten times as strong as the other, this equation gives

$$E = 0.060 \times 10^8 \text{ C.G.S. units} \\ = 0.060 \text{ volts.}$$

W. Nernst, to whom this theory is due, determined the electromotive force of this cell experimentally, and found the value 0.055 volt.

The logarithmic formulae for these concentration cells indicate that theoretically their electromotive force can be increased to any extent by diminishing without limit the concentration of the more dilute solution, log c_1/c_2 then becoming very great. This condition may be realized to some extent in a manner that throws light on the general theory of the voltaic cell. Let us consider the arrangement—silver | silver chloride with potassium chloride solution | potassium nitrate solution | silver nitrate solution | silver. Silver chloride is a very insoluble substance, and here the amount in solution is still further reduced by the presence of excess of chlorine ions of the potassium salt. Thus silver, at one end of the cell in contact with many silver ions of the silver nitrate solution, at the other end is in contact with a liquid in which the concentration of those ions is very small indeed. The result is that a high electromotive force is set up, which has been calculated as 0.52 volt, and observed as 0.51 volt. Again, Hittorf has shown that the effect of a cyanide round a copper electrode is to combine with the copper ions. The concentration of the simple copper ions is then so much diminished that the copper plate becomes an anode with regard to zinc. Thus the cell—copper | potassium cyanide solution | potassium sulphate solution—zinc sulphate solution | zinc—gives a current which carries copper into solution and deposits zinc. In a similar way silver could be made to act as anode with respect to cadmium.

It is now evident that the electromotive force of an ordinary chemical cell such as that of Daniell depends on the concentration of the solutions as well as on the nature of the metals. In ordinary cases possible changes in the concentrations only affect the electromotive force by a few parts in a hundred, but, by means such as those indicated above, it is possible to produce such immense differences in the concentrations that the electromotive force of the cell is not only changed appreciably but even reversed in direction. Once more we see that it is the total impending change in the available energy of the system which controls the electromotive force.

Any reversible cell can theoretically be employed as an accumulator, though, in practice, conditions of general convenience are more sought after than thermodynamic efficiency. The effective electromotive force of the common lead accumulator (*q.v.*) is less than that required to charge it. This drop in the electromotive force has led to the belief that the cell is not reversible. F. Dolezalek, however, has attributed the difference to mechanical hindrances, which prevent the equalization of acid concentration in the neighbourhood of the electrodes, rather than to any essentially irreversible chemical action. The fact that the Gibbs-Helmholtz equation is found to apply also indicates that the lead accumulator is approximately reversible in the thermodynamic sense of the term.

Polarization and Contact Difference of Potential.—If we connect together in series a single Daniell's cell, a galvanometer, and two platinum electrodes dipping into acidulated water, no visible chemical decomposition ensues. At first a considerable current is indicated by the galvanometer; the deflexion soon diminishes, however, and finally becomes very small. If, instead of using a single Daniell's cell, we employ some source of electromotive force which can be varied as we please, and gradually raise its intensity, we shall find that, when it exceeds a certain value, about 1.7 volt, a permanent current of considerable strength flows through the solution, and, after the initial period, shows no signs of decrease. This current is accompanied by chemical decomposition. Now let us disconnect the platinum plates from the battery and join them directly with the galvanometer. A current will flow for a while in the reverse direction; the system of plates and acidulated water through which a current has been passed, acts as an accumulator, and will itself yield a current in return. These phenomena are explained by the existence of a reverse electromotive force at the surface of the platinum plates. Only when the applied electromotive force exceeds this reverse force of polarization, will a permanent steady current pass through the liquid, and visible chemical decomposition proceed. It seems that this reverse electromotive force of polarization is due to the deposit on the electrodes of minute quantities of the products of chemical decomposition. Differences between the two electrodes are thus set up, and, as we have seen above, an electromotive force will therefore exist between them. To pass a steady current in the direction opposite to this electromotive force of polarization, the applied electromotive force E must exceed that of polarization E′, and the excess E − E′ is the effective electromotive force of the circuit, the current being, in accordance with Ohm's law, proportional to the applied electromotive force and represented by (E − E′) / R, where R is a constant called the resistance of the circuit.

When we use platinum electrodes in acidulated water, hydrogen and oxygen are evolved. The opposing force of polarization is about 1.7 volt, but, when the plates are disconnected and used as a source of current, the electromotive force they give is only about 1.07 volt. This irreversibility is due to the work required to evolve bubbles of gas at the surface of bright platinum plates. If the plates be covered with a deposit of platinum black, in which the gases are absorbed as fast as they are produced, the minimum decomposition point is 1.07 volt, and the process is reversible. If secondary effects are eliminated, the deposition of metals also is a reversible process; the decomposition voltage is equal to the electromotive force which the metal itself gives when going into solution. The phenomena of polarization are thus seen to be due to the changes of surface produced, and are correlated with the differences of potential which exist at any surface of separation between a metal and an electrolyte.

Many experiments have been made with a view of separating the two potential-differences which must exist in any cell made of two metals and a liquid, and of determining each one individually. If we regard the thermal effect at each junction as a measure of the potential-difference there, as the total thermal effect in the cell undoubtedly is of the sum of its potential-differences, in cases where the temperature coefficient is negligible, the heat evolved on solution of a metal should give the electrical potential-difference at its surface. Hence, if we assume that, in the Daniell's cell, the temperature coefficients are negligible at the individual contacts as well as in the cell as a whole, the sign of the potential-difference ought to be the same at the surface of the zinc as it is at the surface of the copper. Since zinc goes into solution and copper comes out, the electromotive force of the cell will be the difference between the two effects. On the other hand, it is commonly thought that the single potential-differences at the surface of metals and electrolytes have been determined by methods based on the use of the capillary electrometer and on others depending on what is called a dropping electrode, that is, mercury dropping rapidly into an electrolyte and forming a cell with the mercury at rest in the bottom of the vessel. By both these methods the single potential-differences found at the surfaces of the zinc and copper have opposite signs, and the effective electromotive force of a Daniell's cell is the sum of the two effects. Which of these conflicting views represents the truth still remains uncertain.

Diffusion of Electrolytes and Contact Difference of Potential between Liquids.—An application of the theory of ionic velocity due to W. Nernst[7] and M. Planck[8] enables us to calculate the diffusion constant of dissolved electrolytes. According to the molecular theory, diffusion is due to the motion of the molecules of the dissolved substance through the liquid. When the dissolved molecules are uniformly distributed, the osmotic pressure will be the same everywhere throughout the solution, but, if the concentration vary from point to point, the pressure will vary also. There must, then, be a relation between the rate of change of the concentration and the osmotic pressure gradient, and thus we may consider the osmotic pressure gradient as a force driving the solute through a viscous medium. In the case of non-electrolytes and of all non-ionized molecules this analogy completely represents the facts, and the phenomena of diffusion can be deduced from it alone. But the ions of an electrolytic solution can move independently through the liquid, even when no current flows, as the consequences of Ohm's law indicate. The ions will therefore diffuse independently, and the faster ion will travel quicker into pure water in contact with a solution. The ions carry their charges with them, and, as a matter of fact, it is found that water in contact with a solution takes with respect to it a positive or negative potential, according as the positive or negative ion travels the faster. This process will go on until the simultaneous separation of electric charges produces an electrostatic force strong enough to prevent further separation of ions. We can therefore calculate the rate at which the salt as a whole will diffuse by examining the conditions for a steady transfer, in which the ions diffuse at an equal rate, the faster one being restrained and the slower one urged forward by the electric forces. In this manner the diffusion constant can be calculated in absolute units (HCl = 2.49, HNO_3 = 2.27, $NaCl$ = 1.12), the unit of time being the day. By experiments on diffusion this constant has been found by Scheffer, and the numbers observed agree with those calculated (HCl = 2.30, HNO_3 = 2.22, $NaCl$ = 1.11).

As we have seen above, when a solution is placed in contact with water the water will take a positive or negative potential with regard to the solution, according as the cation or anion has the greater specific velocity, and therefore the greater initial rate of diffusion. The difference of potential between two solutions of a substance at different concentrations can be calculated from the equations used to give the diffusion constants. The results give equations of the same logarithmic form as those obtained in a somewhat different manner in the theory of concentration cells described above, and have been verified by experiment.

The contact differences of potential at the interfaces of metals and electrolytes have been co-ordinated by Nernst with those at the surfaces of separation between different liquids. In contact with a solvent a metal is supposed to possess a definite solution pressure, analogous to the vapour pressure of a liquid. Metal goes into solution in the form of electrified ions. The liquid thus acquires a positive charge, and the metal a negative charge. The electric forces set up tend to prevent further separation, and finally a state of equilibrium is reached, when no more ions can go into solution unless an equivalent number are removed by 226
voltaic action. On the analogy between this case and that of the interface between two solutions, Nernst has arrived at similar logarithmic expressions for the difference of potential, which becomes proportional to log (P_1/P_2) where P_2 is taken to mean the osmotic pressure of the cations in the solution, and P_1 the osmotic pressure of the cations in the substance of the metal itself. On these lines the equations of concentration cells, deduced above on less hypothetical grounds, may be regained.

Theory of Electrons.—Our views of the nature of the ions of electrolytes have been extended by the application of the ideas of the relations between matter and electricity obtained by the study of electric conduction through gases. The interpretation of the phenomena of gaseous conduction was rendered possible by the knowledge previously acquired of conduction through liquids; the newer subject is now reaching a position whence it can repay its debt to the older.

Sir J.J. Thomson has shown (see Conduction, Electric, § III.) that the negative ions in certain cases of gaseous conduction are much more mobile than the corresponding positive ions, and possess a mass of about the one-thousandth part of that of a hydrogen atom. These negative particles or corpuscles seem to be the ultimate units of negative electricity, and may be identified with the electrons required by the theories of H.A. Lorentz and Sir J. Larmor. A body containing an excess of these particles is negatively electrified, and is positively electrified if it has parted with some of its normal number. An electric current consists of a moving stream of electrons. In gases the electrons sometimes travel alone, but in liquids they are always attached to matter, and their motion involves the movement of chemical atoms or groups of atoms. An atom with an extra corpuscle is a univalent negative ion, an atom with one corpuscle detached is a univalent positive ion. In metals the electrons can slip from one atom to the next, since a current can pass without chemical action. When a current passes from an electrolyte to a metal, the electron must be detached from the atom it was accompanying and chemical action be manifested at the electrode.

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Some of the more important papers on the subject have been reprinted for Harper's *Series of Scientific Memoirs in Electrolytic Conduction* (1899) and the *Modern Theory of Solution* (1899). Several journals are published specially to deal with physical chemistry, of which electrochemistry forms an important part. Among them may be mentioned the *Zeitschrift für physikalische Chemie* (Leipzig); and the *Journal of Physical Chemistry* (Cornell University). In these periodicals will be found new work on the subject and abstracts of papers which appear in other physical and chemical publications.

(W. C. D. W.)

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1 See Hittorf, *Pogg. Ann.* cvi. 517 (1859).

2 *Grundriss der Elektrochemie* (1895), p. 292; see also F. Kaufler and C. Herzog, *Ber.*, 1909, 42, p. 3858.

3 *Brit. Ass. Rep.*, 1906, Section A, Presidential Address.

4 See *Theory of Solution*, by W.C.D. Whetham (1902), p. 328.

5 W. Ostwald, *Zeits. physikal. Chemie*, 1892, vol. IX. p. 579; T. Ewan, *Phil. Mag.* (5), 1892, vol. xxxiii. p. 317; G.D. Liveing, *Cambridge Phil. Trans.*, 1900, vol. xviii. p. 298.

6 See W.B. Hardy, *Journal of Physiology*, 1899, vol. xxiv. p. 288; and W.C.D. Whetham, *Phil. Mag.*, November 1899.

7 *Zeits. physikal. Chem.* 2, p. 613.

8 *Wied. Ann.*, 1890, 40, p. 561.

ELECTROMAGNETISM, that branch of physical science which is concerned with the interconnexion of electricity and magnetism, and with the production of magnetism by means of electric currents by devices called electromagnets.

History.—The foundation was laid by the observation first made by Hans Christian Oersted (1777-1851), professor of natural philosophy in Copenhagen, who discovered in 1820 that a wire uniting the poles or terminal plates of a voltaic pile has the property of affecting a magnetic needle[1] (see ELECTRICITY). Oersted carefully ascertained that the nature of the wire itself did not influence the result but saw that it was due to the electric conflict, as he called it, round the wire; or in modern language, to the magnetic force or magnetic flux round the conductor. If a straight wire through which an electric current is flowing is placed above and parallel to a magnetic compass needle, it is found that if the current is flowing in the conductor in a direction from south to north, the north pole of the needle under the conductor deviates to the left hand, whereas if the conductor is placed under the needle, the north pole deviates to the right hand; if the conductor is doubled back over the needle, the effects of the two sides of the loop are added together and the deflection is increased. These results are summed up in the mnemonic rule: *Imagine yourself swimming in the conductor with the current, that is, moving in the direction of the positive electricity, with your face towards the magnetic needle; the north pole will then deviate to your left hand.* The deflection of the magnetic needle can therefore reveal the existence of an electric current in a neighbouring circuit, and this fact was soon utilized in the construction of instruments called galvanometers (*q.v.*).

Immediately after Oersted's discovery was announced, D.F.J. Arago and A.M. Ampère began investigations on the subject of electromagnetism. On the 18th of September 1820, Ampère read a paper before the Academy of Sciences in Paris, in which he announced that the voltaic pile itself affected a magnetic needle as did the uniting wire, and he showed that the effects in both cases were consistent with the theory that electric current was a circulation round a circuit, and equivalent in magnetic effect to a very short magnet with axis placed at right angles to the plane of the circuit. He then propounded his brilliant hypothesis that the magnetization of iron was due to molecular electric currents. This suggested to Arago that wire wound into a helix carrying electric current should magnetize a steel needle placed in the interior. In the *Ann. Chim.* (1820, 15, p. 94), Arago published a paper entitled "Expériences relatives à l'aimantation du fer et de l'acier par l'action du courant voltaïque," announcing that the wire conveying the current, even though of copper, could magnetize steel needles placed across it, and if plunged into iron filings it attracted them. About the same time Sir Humphry Davy sent a communication to Dr W.H. Wollaston, read at the Royal Society on the 16th of November 1820 (reproduced in the *Annals of Philosophy* for August 1821, p. 81), "On the Magnetic Phenomena produced by Electricity," in which he announced his independent discovery of the same fact. With a large battery of 100 pairs of plates at the Royal Institution, he found in October 1820 that the uniting wire became strongly magnetic and that iron filings clung to it; also that steel needles placed across the wire were permanently magnetized. He placed a sheet of glass over the wire and sprinkling iron filings on it saw that they arranged themselves in straight lines at right angles to the wire. He then proved that Leyden jar discharges could produce the same effects. Ampère and Arago then seem to have experimented together and magnetized a steel needle wrapped in paper which was enclosed in a helical wire conveying a current. All these facts were rendered intelligible when it was seen that a wire when conveying an electric current becomes surrounded by a magnetic field. If the wire is a long straight one, the lines of magnetic force are circular and concentric with centres on the wire axis, and if the wire is bent into a circle the lines of magnetic force are endless loops surrounding and linked with the electric circuit. Since a magnetic pole tends to move along a line of magnetic force it was obvious that it should revolve round a wire conveying a current. To exhibit this fact involved, however, much ingenuity. It was first accomplished by Faraday in October 1821 (*Exper. Res.* ii. p. 127). Since the action is reciprocal a current free to move tends to revolve round a magnetic pole. The fact is most easily shown by a small piece of apparatus made as follows: In a glass cylinder (see fig. 1) like a lamp chimney are fitted two corks. Through the bottom one is passed the north end of a bar magnet which projects up above a little mercury lying in the cork. Through the top cork is passed one end of a wire from a battery, and a piece of wire in the cylinder is flexibly connected to it, the 227
lower end of this last piece just touching the mercury. When a current is passed in at the top wire and out at the lower end of the bar magnet, the loose wire revolves round the magnet pole. All text-books on physics contain in their chapters on electromagnetism full accounts of various forms of this experiment.

In 1825 another important step forward was taken when William Sturgeon (1783-1850) of London produced the electromagnet. It consisted of a horseshoe-shaped bar of soft iron, coated with varnish, on which was wrapped a spiral coil of bare copper wire, the turns not touching each other. When a voltaic current was passed through the wire the iron became a powerful magnet, but on severing the connexion with the battery, the soft iron lost immediately nearly all its magnetism.[2]

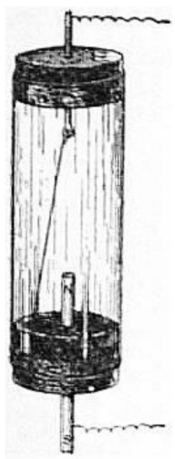

FIG. 1.

At that date Ohm had not announced his law of the electric circuit, and it was a matter of some surprise to investigators to find that Sturgeon's electromagnet could not be operated at a distance through a long circuit of wire with such good results as when close to the battery. Peter Barlow, in January 1825, published in the *Edinburgh Philosophical Journal*, a description of such an experiment made with a view of applying Sturgeon's electromagnet to telegraphy, with results which were unfavourable. Sturgeon's experiments, however, stimulated Joseph Henry (*q.v.*) in the United States, and in 1831 he gave a description of a method of winding electromagnets which at once put a new face upon matters (*Silliman's Journal*, 1831, 19, p. 400). Instead of insulating the iron core, he wrapped the copper wire round with silk and wound in numerous turns and many layers upon the iron horseshoe in such fashion that the current went round the iron always in the same direction. He then found that such an electromagnet wound with a long fine wire, if worked with a battery consisting of a large number of cells in series, could be operated at a considerable distance, and he thus produced what were called at that time *intensity electromagnets*, and which subsequently rendered the electric telegraph a possibility. In fact, Henry established in 1831, in Albany, U.S.A., an electromagnetic telegraph, and in 1835 at Princeton even used an earth return, thereby anticipating the discovery (1838) of C.A. Steinheil (1801-1870) of Munich.

Inventors were then incited to construct powerful electromagnets as tested by the weight they could carry from their armatures. Joseph

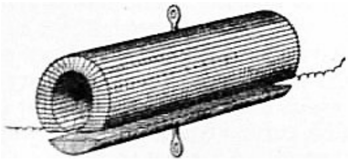

FIG. 2.

Henry made a magnet for Yale College, U.S.A., which lifted 3000 ℔ (*Silliman's Journal*, 1831, 20, p. 201), and one for Princeton which lifted 3000 with a very small battery. Amongst others J.P. Joule, ever memorable for his investigations on the mechanical equivalent of heat, gave much attention about 1838-1840 to the construction of electromagnets and succeeded in devising some forms remarkable for their lifting power. One form was constructed by cutting a thick soft iron tube longitudinally into two equal parts. Insulated copper wire was then wound longitudinally over one of both parts (see fig. 2) and a current sent through the wire. In another form two iron disks with teeth at right angles to the disk had insulated wire wound zigzag between the teeth; when a current was sent through the wire, the teeth were so magnetized that they were alternately N. and S. poles. If two such similar disks were placed with teeth of opposite polarity in contact, a very large force was required to detach them, and with a magnet and armature weighing in all 11.575 ℔ Joule found that a weight of 2718 was supported. Joule's papers on this subject will be found in his *Collected Papers* published by the Physical Society of London, and in *Sturgeon's Annals of Electricity*, 1838-1841, vols. 2-6.

The Magnetic Circuit.—The phenomena presented by the electromagnet are interpreted by the aid of the notion of the magnetic circuit. Let us consider a thin circular sectioned ring of iron wire wound over with a solenoid or spiral of insulated copper wire through which a current of electricity can be passed. If the solenoid or wire windings existed alone, a current having a strength A amperes passed through it would create in the interior of the solenoid a magnetic force H, numerically equal to $4\pi/10$ multiplied by the number of windings N on the solenoid, and by the current in amperes A, and divided by the mean length of the solenoid l, or $H = 4\pi AN/10l$. The product AN is called the "ampere-turns" on the solenoid. The product Hl of the magnetic force H and the length l of the magnetic circuit is called the "magnetomotive force" in the magnetic circuit, and from the above formula it is seen that the magnetomotive force denoted by (M.M.F.) is equal to $4\pi/10$ (= 1.25 nearly) times the ampere-turns (A.N.) on the exciting coil or solenoid. Otherwise (A.N.) = 0.8(M.M.F.). The magnetomotive force is regarded as creating an effect called magnetic flux (Z) in the magnetic circuit, just as electromotive force E.M.F. produces electric current (A) in the electric circuit, and as by Ohm's law (see ELECTROKINETICS) the current varies as the E.M.F. and inversely as a quality of the electric circuit called its "resistance," so in the magnetic circuit the magnetic flux varies as the magnetomotive force and inversely as a quality of the magnetic circuit called its "reluctance." The great difference between the electric circuit and the magnetic circuit lies in the fact that whereas the electric resistance of a solid or liquid conductor is independent of the current and affected only by the temperature, the magnetic reluctance varies with the magnetic flux and cannot be defined except by means of a curve which shows its value for different flux densities. The quotient of the total magnetic flux, Z, in a circuit by the cross section, S, of the circuit is called the mean "flux density," and the reluctance of a magnetic circuit one centimetre long and one square centimetre in cross section is called the "reluctivity" of the material. The relation between reluctivity $\rho = 1/\mu$ magnetic force H, and flux density B, is defined by the equation $H = \rho B$, from which we have $Hl = Z (\rho l/S)$ = M.M.F. acting on the circuit. Again, since the ampere-turns (AN) on the circuit are equal to 0.8 times the M.M.F., we have finally $AN/l = 0.8(Z/\mu S)$. This equation tells us the exciting force reckoned in ampere-turns, AN, which must be put on the ring core to create a total magnetic flux Z in it, the ring core having a mean perimeter l and cross section S and reluctivity $\rho = 1/\mu$ corresponding to a flux density Z/S. Hence before we can make use of the equation for practical purposes we need to possess a curve for the particular material showing us the value of the reluctivity corresponding to various values of the possible flux density. The reciprocal of ρ is usually called the "permeability" of the material and denoted by μ. Curves showing the relation of $1/\rho$ and ZS or μ and B, are called "permeability curves." For air and all other non-magnetic matter the permeability has the same value, taken arbitrarily as unity. On the other hand, for iron, nickel and cobalt the permeability may in some cases reach a value of 2000 or 2500 for a value of B = 5000 in C.G.S. measure (see UNITS, PHYSICAL). The process of taking these curves consists in sending a current of known strength through a solenoid of known number of turns wound on a circular iron ring of known dimensions, and observing the time-integral of the secondary current produced in a secondary circuit of known turns and resistance R wound over the iron core N times. The secondary electromotive force is by Faraday's law (see ELECTROKINETICS) equal to the time rate of change of the total flux, or $E = NdZ/dt$. But by Ohm's law $E = Rdq/dt$, where q is the quantity of electricity set flowing in the secondary circuit by a change dZ in the co-linked total flux. Hence if 2Q represents this total quantity of electricity set flowing in the secondary circuit by suddenly reversing the direction of the magnetic flux Z in the iron core we must have

$$RQ = NZ \text{ or } Z = RQ/N.$$

The measurement of the total quantity of electricity Q can be made by means of a ballistic galvanometer (*q.v.*), and the resistance R of the secondary circuit includes that of the coil wound on the iron core and the galvanometer as well. In this manner the value of the total flux Z and therefore of Z/S = B or the flux density, can be found for a given magnetizing force H, and this last quantity is determined when we know the magnetizing current in the solenoid and its turns and dimensions. The curve which delineates the relation of H and B is called the magnetization curve for the material in question. For examples of these curves see MAGNETISM.

The fundamental law of the non-homogeneous magnetic circuit traversed by one and the same total magnetic flux Z is that the sum of all the magnetomotive forces acting in the circuit is numerically equal to the product of the factor 0.8, the total flux in the circuit, and the sum of all the reluctances of the various parts of the circuit. If then the circuit consists of materials of different permeability and it is desired to know the ampere-turns required to produce a given total of flux round the circuit, we have to calculate from the magnetization curves of the material of each part the necessary magnetomotive forces and add these forces together. The practical application of this principle to the predetermination of the field windings of dynamo magnets was first made by Drs J. and E. Hopkinson (*Phil. Trans.*, 1886, 177, p. 331). 228

We may illustrate the principles of this predetermination by a simple example. Suppose a ring of iron has a mean diameter of 10 cms. and a cross section of 2 sq. cms., and a transverse cut on air gap made in it 1 mm. wide. Let us inquire the ampere-turns to be put upon the ring to create in it a total flux of 24,000 C.G.S. units. The total length of the iron part of the circuit is $(10\pi - 0.1)$ cms., and its section is 2 sq. cms., and the flux density in it is to be 12,000. From Table II. below we see that the permeability of pure iron corresponding to a flux density of 12,000 is 2760. Hence the reluctance of the iron circuits is equal to

$$\frac{10\pi - 0.1}{2760 \times 2} = \frac{220}{38640} \text{ C.G.S. units.}$$

The length of the air gap is 0.1 cm., its section 2 sq. cms., and its permeability is unity. Hence the reluctance of the air gap is

$$\frac{0.1}{1 \times 2} = \frac{1}{20} \text{ C.G.S. unit.}$$

Accordingly the magnetomotive force in ampere-turns required to produce the required flux is equal to

$$0.8\ (24{,}000) \left(\frac{1}{20} + \frac{220}{38640} \right) = 1070 \text{ nearly.}$$

It follows that the part of the magnetomotive force required to overcome the reluctance of the narrow air gap is about nine times that required for the iron alone.

In the above example we have for simplicity assumed that the flux in passing across the air gap does not spread out at all. In dealing with electromagnet design in dynamo construction we have, however, to take into consideration the spreading as well as the leakage of flux across the circuit (see DYNAMO). It will be seen, therefore, that in order that we may predict the effect of a certain kind of iron or steel when used as the core of an electromagnet, we must be provided with tables or curves showing the reluctivity or permeability corresponding to various flux densities or—which comes to the same thing—with (B, H) curves for the sample.

Iron and Steel for Electromagnetic Machinery.—In connexion with the technical application of electromagnets such as those used in the field magnets of dynamos (*q.v.*), the testing of different kinds of iron and steel for magnetic permeability has therefore become very important. Various instruments called permeameters and hysteresis meters have been designed for this purpose, but much of the work has been done by means of a ballistic galvanometer and test ring as above described. The "hysteresis" of an iron or steel is that quality of it in virtue of which energy is dissipated as heat when the magnetization is reversed or carried through a cycle (see MAGNETISM), and it is generally measured either in ergs per cubic centimetre of metal per cycle of magnetization, or in watts per ℔ per 50 or 100 cycles per second at or corresponding to a certain maximum flux density, say 2500 or 600 C.G.S. units. For the details of various forms of permeameter and hysteresis meter technical books must be consulted.[3]

An immense number of observations have been carried out on the magnetic permeability of different kinds of iron and steel, and in the following tables are given some typical results, mostly from experiments made by J.A. Ewing (see *Proc. Inst. C.E.*, 1896, 126, p. 185) in which the ballistic method was employed to determine the flux density corresponding to various magnetizing forces acting upon samples of iron and steel in the form of rings.

The figures under heading I. are values given in a paper by A.W.S. Pocklington and F. Lydall (*Proc. Roy. Soc.*, 1892-1893, 52, pp. 164 and 228) as the results of a magnetic test of an exceptionally pure iron supplied for the purpose of experiment by Colonel Dyer, of the Elswick Works. The substances other than iron in this sample were stated to be: carbon, *trace*; silicon, *trace*; phosphorus, *none*; sulphur, 0.013%; manganese, 0.1%. The other five specimens, II. to VI., are samples of commercial iron or steel. No. II. is a sample of Low Moor bar iron forged into a ring, annealed and turned. No. III. is a steel forging furnished by Mr R. Jenkins as a sample of forged ingot-metal for dynamo magnets. No. IV. is a steel casting for dynamo magnets, unforged, made by Messrs Edgar Allen & Company by a special pneumatic process under the patents of Mr A. Tropenas. No. V. is also an unforged steel casting for dynamo magnets, made by Messrs Samuel Osborne & Company by the Siemens process. No. VI. is also an unforged steel casting for dynamo magnets, made by Messrs Fried. Krupp, of Essen.

TABLE I.—*Magnetic Flux Density corresponding to various Magnetizing Forces in the case of certain Samples of Iron and Steel* (*Ewing*).

Magnetizing Force H (C.G.S. Units).	Magnetic Flux Density B (C.G.S. Units).					
	I.	II.	III.	IV.	V.	VI.
5	12,700	10,900	12,300	4,700	9,600	10,900
10	14,980	13,120	14,920	12,250	13,050	13,320
15	15,800	14,010	15,800	14,000	14,600	14,350
20	16,300	14,580	16,280	15,050	15,310	14,950
30	16,950	15,280	16,810	16,200	16,000	15,660
40	17,350	15,760	17,190	16,800	16,510	16,150
50	. .	16,060	17,500	17,140	16,900	16,480
60	. .	16,340	17,750	17,450	17,180	16,780
70	. .	16,580	17,970	17,750	17,400	17,000
80	. .	16,800	18,180	18,040	17,620	17,200
90	. .	17,000	18,390	18,230	17,830	17,400
100	. .	17,200	18,600	18,420	18,030	17,600

It will be seen from the figures and the description of the materials that the steel forgings and castings have a remarkably high permeability under small magnetizing force.

Table II. shows the magnetic qualities of some of these materials as found by Ewing when tested with small magnetizing forces.

Table II.—*Magnetic Permeability of Samples of Iron and Steel under Weak Magnetizing Forces.*

Magnetic Flux Density B (C.G.S. Units).	I. Pure Iron.		III. Steel Forging.		VI. Steel Casting.	
	H	μ	H	μ	H	μ
2,000	0.90	2220	1.38	1450	1.18	1690
4,000	1.40	2850	1.91	2090	1.66	2410
6,000	1.85	3240	2.38	2520	2.15	2790
8,000	2.30	3480	2.92	2740	2.83	2830
10,000	3.10	3220	3.62	2760	4.05	2470
12,000	4.40	2760	4.80	2500	6.65	1810

The numbers I., III. and VI. in the above table refer to the samples mentioned in connexion with Table I.

It is a remarkable fact that certain varieties of low carbon steel (commonly called mild steel) have a higher permeability than even annealed Swedish wrought iron under large magnetizing forces. The term *steel*, however, here used has reference rather to the mode of production than the final chemical nature of the material. In some of the mild-steel castings used for dynamo electromagnets it appears that the total foreign matter, including carbon, manganese and silicon, is not more than 0.3% of the whole, the material being 99.7% pure iron. This valuable magnetic property of steel capable of being cast is, however, of great utility in modern dynamo building, as it enables field magnets of very high permeability to be constructed, which can be fashioned into shape by casting instead of being built up as formerly out of masses of forged wrought iron. The curves in fig. 3 illustrate the manner in which the flux density or, as it is usually called, the magnetization curve of this mild cast steel crosses that of Swedish wrought iron, and enables us to obtain a higher flux density corresponding to a given magnetizing force with the steel than with the iron.

From the same paper by Ewing we extract a number of results relating to permeability tests of thin sheet iron and sheet steel, such as is used in the construction of dynamo armatures and transformer cores.

No. VII. is a specimen of good transformer-plate, 0.301 millimetre thick, rolled from Swedish iron by Messrs Sankey of Bilston. No. VIII. is a specimen of specially thin transformer-plate rolled from scrap iron. No. IX. is a specimen of transformer-plate rolled from ingot-steel. No. X. is a specimen of the wire which was used by J. Swinburne to form the core of his “hedgehog” transformers. Its diameter was 0.602 millimetre. All these samples were tested in the form of rings by the ballistic method, the rings of sheet-metal being stamped or turned in the flat. The wire ring No. X. was coiled and annealed after coiling.

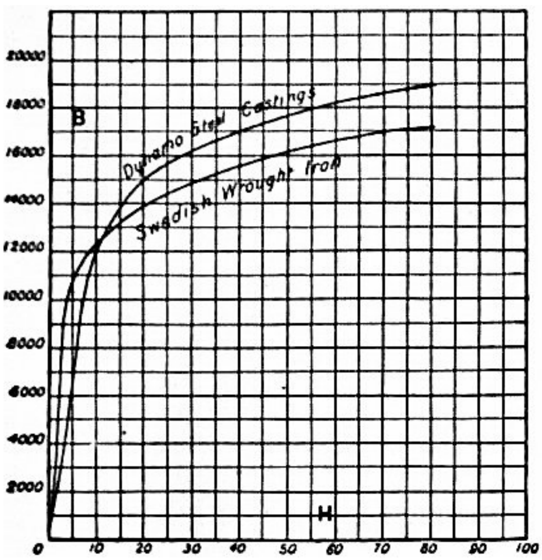

Fig. 3.

Table III.—*Permeability Tests of Transformer Plate and Wire.*

Magnetic Flux Density B (C.G.S. Units).	VII. Transformer-plate of Swedish Iron.		VIII. Transformer-plate of Scrap Iron.		IX. Transformer-plate of Steel.		X. Transformer-wire.	
	H	μ	H	μ	H	μ	H	μ
1,000	0.81	1230	1.08	920	0.60	1470	1.71	590
2,000	1.05	1900	1.46	1370	0.90	2230	2.10	950
3,000	1.26	2320	1.77	1690	1.04	2880	2.30	1300
4,000	1.54	2600	2.10	1900	1.19	3360	2.50	1600
5,000	1.82	2750	2.53	1980	1.38	3620	2.70	1850
6,000	2.14	2800	3.04	1970	1.59	3770	2.92	2070
7,000	2.54	2760	3.62	1930	1.89	3700	3.16	2210
8,000	3.09	2590	4.37	1830	2.25	3600	3.43	2330
9,000	3.77	2390	5.3	1700	2.72	3310	3.77	2390
10,000	4.6	2170	6.5	1540	3.33	3000	4.17	2400

11,000	5.7	1930	7.9	1390	4.15	2650	4.70	2340
12,000	7.0	1710	9.8	1220	5.40	2220	5.45	2200
13,000	8.5	1530	11.9	1190	7.1	1830	6.5	2000
14,000	11.0	1270	15.0	930	10.0	1400	8.4	1670
15,000	15.1	990	19.5	770	· ·	· ·	11.9	1260
16,000	21.4	750	27.5	580	· ·	· ·	21.0	760

Some typical flux-density curves of iron and steel as used in dynamo and transformer building are given in fig. 4.

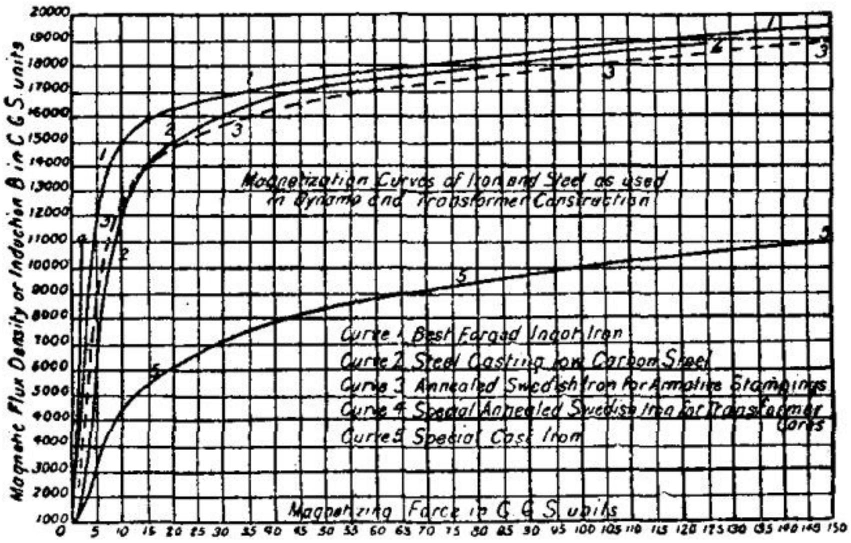

Fig. 4.

The numbers in Table III. well illustrate the fact that the permeability, μ = B/H has a maximum value corresponding to a certain flux density. The tables are also explanatory of the fact that mild steel has gradually replaced iron in the manufacture of dynamo electromagnets and transformer-cores.

Broadly speaking, the materials which are now employed in the manufacture of the cores of electromagnets for technical purposes of various kinds may be said to fall into three classes, namely, forgings, castings and stampings. In some cases the iron or steel core which is to be magnetized is simply a mass of iron hammered or pressed into shape by hydraulic pressure; in other cases it has to be fused and cast; and for certain other purposes it must be rolled first into thin sheets, which are subsequently stamped out into the required forms.

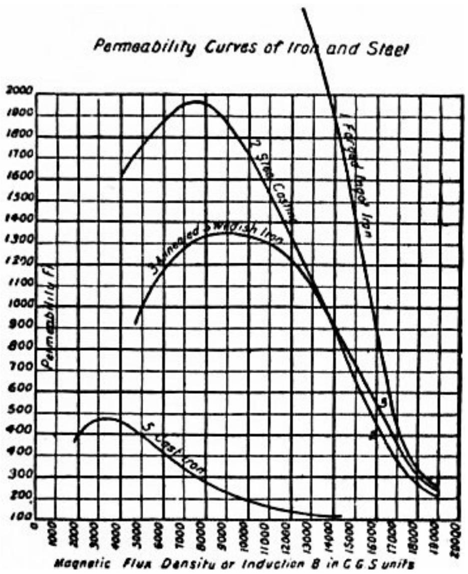

Fig. 5.

For particular purposes it is necessary to obtain the highest possible magnetic permeability corresponding to a high, or the highest attainable flux density. This is generally the case in the electromagnets which are employed as the field magnets in dynamo machines. It may generally be said that whilst the best wrought iron, such as annealed Low Moor or Swedish iron, is more permeable for low flux densities than steel castings, the cast steel may surpass the wrought metal for high flux density. For most electro-technical purposes the best magnetic results are given by the employment of forged ingot-iron. This material is probably the most permeable throughout the whole scale of attainable flux densities. It is slightly superior to wrought iron, and it only becomes inferior to the highest class of cast steel when the flux density is pressed above 18,000 C.G.S. units (see fig. 5). For flux densities above 13,000 the forged ingot-iron has now practically replaced for electric engineering purposes the Low Moor or Swedish iron. Owing to the method of its production, it might in truth be called a soft steel with a very small percentage of combined carbon. The

best description of this material is conveyed by the German term “Flusseisen,” but its nearest British equivalent is “ingot-iron.” Chemically speaking, the material is for all practical purposes very nearly pure iron. The same may be said of the cast steels now much employed for the production of dynamo magnet cores. The cast steel which is in demand for this purpose has a slightly lower permeability than the ingot-iron for low flux densities, but for flux densities above 16,000 the required result may be more cheaply obtained with a steel casting than with a forging. When high tensile strength is required in addition to considerable magnetic permeability, it has been found advantageous to employ a steel containing 5% of nickel. The rolled sheet iron and sheet steel which is in request for the construction of magnet cores, especially those in which the exciting current is an alternating current, are, generally speaking, produced from Swedish iron. Owing to the mechanical treatment necessary to reduce the material to a thin sheet, the permeability at low flux densities is rather higher than, although at high flux densities it is inferior to, the same iron and steel when 230
tested in bulk. For most purposes, however, where a laminated iron magnet core is required, the flux density is not pressed up above 6000 units, and it is then more important to secure small hysteresis loss than high permeability. The magnetic permeability of cast iron is much inferior to that of wrought or ingot-iron, or the mild steels taken at the same flux densities.

The following Table IV. gives the flux density and permeability of a typical cast iron taken by J.A. Fleming by the ballistic method:—

Table IV.—*Magnetic Permeability and Magnetization Curve of Cast Iron.*

H	B	μ	H	B	μ	H	B	μ
.19	27	139	8.84	4030	456	44.65	8,071	181
.41	62	150	10.60	4491	424	56.57	8,548	151
1.11	206	176	12.33	4884	396	71.98	9,097	126
2.53	768	303	13.95	5276	378	88.99	9,600	108
3.41	1251	367	15.61	5504	353	106.35	10,066	95
4.45	1898	427	18.21	5829	320	120.60	10,375	86
5.67	2589	456	26.37	6814	258	140.37	10,725	76
7.16	3350	468	36.54	7580	207	152.73	10,985	72

The metal of which the tests are given in Table IV. contained 2% of silicon, 2.85% of total carbon, and 0.5% of manganese. It will be seen that a magnetizing force of about 5 C.G.S. units is sufficient to impart to a wrought-iron ring a flux density of 18,000 C.G.S. units, but the same force hardly produces more than one-tenth of this flux density in cast iron.

The testing of sheet iron and steel for magnetic hysteresis loss has developed into an important factory process, giving as it does a means of ascertaining the suitability of the metal for use in the manufacture of transformers and cores of alternating-current electromagnets.

In Table V. are given the results of hysteresis tests by Ewing on samples of commercial sheet iron and steel. The numbers VII., VIII., IX. and X. refer to the same samples as those for which permeability results are given in Table III.

Table V.—*Hysteresis Loss in Transformer-iron.*

Maximum Flux Density B.	Ergs per Cubic Centimetre per Cycle.				Watts per ℔ at a Frequency of 100.			
	VII. Swedish Iron.	VIII. Forged Scrap-iron.	IX. Ingot-steel.	X. Soft Iron Wire.	VII.	VIII.	IX.	X.
2000	240	400	215	600	0.141	0.236	0.127	0.356
3000	520	790	430	1150	0.306	0.465	0.253	0.630
4000	830	1220	700	1780	0.490	0.720	0.410	1.050
5000	1190	1710	1000	2640	0.700	1.010	0.590	1.550
6000	1600	2260	1350	3360	0.940	1.330	0.790	1.980
7000	2020	2940	1730	4300	1.200	1.730	1.020	2.530
8000	2510	3710	2150	5300	1.480	2.180	1.270	3.120
9000	3050	4560	2620	6380	1.800	2.680	1.540	3.750

In Table VI. are given the results of a magnetic test of some exceedingly good transformer-sheet rolled from Swedish iron.

Table VI.—*Hysteresis Loss in Strip of Transformer-plate rolled Swedish Iron.*

Maximum Flux Density B.	Ergs per Cubic Centimetre per Cycle.	Watts per ℔ at a Frequency of 100.
2000	220	0.129
3000	410	0.242
4000	640	0.376
5000	910	0.535
6000	1200	0.710
7000	1520	0.890
8000	1900	1.120
9000	2310	1.360

In Table VII. are given some values obtained by Fleming for the hysteresis loss in the sample of cast iron, the permeability test of which is recorded in Table IV.

Table VII.—*Observations on the Magnetic Hysteresis of Cast Iron.*

Loop.	B (max.)	Hysteresis Loss.	
		Ergs per cc. per Cycle.	Watts per ℔ per. 100 Cycles per sec.
I.	1475	466	.300
II.	2545	1,288	.829
III.	3865	2,997	1.934
IV.	5972	7,397	4.765
V.	8930	13,423	8.658

For most practical purposes the constructor of electromagnetic machinery requires his iron or steel to have some one of the following characteristics. If for dynamo or magnet making, it should have the highest possible permeability at a flux density corresponding to practically maximum magnetization. If for transformer or alternating-current magnet building, it should have the smallest possible hysteresis loss at a maximum flux density of 2500 C.G.S. units during the cycle. If required for permanent magnet making, it should have the highest possible coercivity combined with a high retentivity. Manufacturers of iron and steel are now able to meet these demands in a very remarkable manner by the commercial production of material of a quality which at one time would have been considered a scientific curiosity.

It is usual to specify iron and steel for the first purpose by naming the minimum permeability it should possess corresponding to a flux density of 18,000 C.G.S. units; for the second, by stating the hysteresis loss in watts per ℔ per 100 cycles per second, corresponding to a maximum flux density of 2500 C.G.S. units during the cycle; and for the third, by mentioning the coercive force required to reduce to zero magnetization a sample of the metal in the form of a long bar magnetized to a stated magnetization. In the cyclical reversal of magnetization of iron we have two modes to consider. In the first case, which is that of the core of the alternating transformer, the magnetic force passes through a cycle of values, the iron remaining stationary, and the direction of the magnetic force being always the same. In the other case, that of the dynamo armature core, the direction of the magnetic force in the iron is constantly changing, and at the same time undergoing a change in magnitude.

It has been shown by F.G. Baily (*Proc. Roy. Soc.*, 1896) that if a mass of laminated iron is rotating in a magnetic field which remains constant in direction and magnitude in any one experiment, the hysteresis loss rises to a maximum as the magnitude of the flux density in the iron is increased and then falls away again to nearly zero value. These observations have been confirmed by other observers. The question has been much debated whether the values of the hysteresis loss obtained by these two different methods are identical for magnetic cycles in which the flux density reaches the same maximum value. This question is also connected with another one, namely, whether the hysteresis loss per cycle is or is not a function of the speed with which the cycle is traversed. Early experiments by C.P. Steinmetz and others seemed to show that there was a difference between slow-speed and high-speed hysteresis cycles, but later experiments by J. Hopkinson and by A. Tanakadaté, though not absolutely exhaustive, tend to prove that up to 400 cycles per second the hysteresis loss per cycle is practically unchanged.

Experiments made in 1896 by R. Beattie and R.C. Clinker on magnetic hysteresis in rotating fields were partly directed to determine whether the hysteresis loss at moderate flux densities, such as are employed in transformer work, was the same as that found by measurements made with alternating-current fields on the same iron and steel specimens (see *The Electrician*, 1896, 37, p. 723). These experiments showed that over 231 moderate ranges of induction, such as may be expected in electro-technical work, the hysteresis loss per cycle per cubic centimetre was practically the same when the iron was tested in an alternating field with a periodicity of 100, the field remaining constant in direction, and when the iron was tested in a rotating field giving the same maximum flux density.

With respect to the variation of hysteresis loss in magnetic cycles having different maximum values for the flux density, Steinmetz found that the hysteresis loss (W), as measured by the area of the complete (B, H) cycle and expressed in ergs per centimetre-cube per cycle, varies proportionately to a constant called the *hysteretic constant*, and to the 1.6th power of the maximum flux density (B), or $W = \eta B^{1.6}$.

The hysteretic constants (η) for various kinds of iron and steel are given in the table below:—

Metal.	Hysteretic Constant.
Swedish wrought iron, well annealed	.0010 to .0017
Annealed cast steel of good quality; small percentage of carbon	.0017 to .0029
Cast Siemens-Martin steel	.0019 to .0028
Cast ingot-iron	.0021 to .0026
Cast steel, with higher percentages of carbon, or inferior qualities of wrought iron	.0031 to .0054

Steinmetz's law, though not strictly true for very low or very high maximum flux densities, is yet a convenient empirical rule for obtaining approximately the hysteresis loss at any one maximum flux density and knowing it at another, provided these values fall within a range varying say from 1 to 9000 C.G.S. units. (See Magnetism.)

The standard maximum flux density which is adopted in electro-technical work is 2500, hence in the construction of the cores of alternating-current electromagnets and transformers iron has to be employed having a known hysteretic constant at the standard flux density. It is generally expressed by stating the number of watts per ℔ of metal which would be dissipated for a frequency of 100 cycles, and a maximum flux density (B max.) during the cycle of 2500. In the case of good iron or steel for transformer-core making, it should not exceed 1.25 watt per ℔ per 100 cycles per 2500 B (maximum value).

It has been found that if the sheet iron employed for cores of alternating electromagnets or transformers is heated to a temperature somewhere in the neighbourhood of 200° C. the hysteresis loss is very greatly increased. It was noticed in 1894 by G.W. Partridge that alternating-current transformers which had been in use some time had a very considerably augmented core loss when compared with their initial condition. O.T. Bláthy and W.M. Mordey in 1895 showed that this augmentation in hysteresis loss in iron was due to heating. H.F. Parshall investigated the effect up to moderate temperatures, such as 140° C., and an extensive series of experiments was made in 1898 by S.R. Roget (*Proc. Roy. Soc.*, 1898, 63, p. 258, and 64, p. 150). Roget found that below 40° C. a rise in temperature did not produce any augmentation in the hysteresis loss in iron, but if it is heated to between 40° C. and 135° C. the hysteresis loss increases continuously with time, and this increase is now called “ageing” of the iron. It proceeds more slowly as the temperature is higher. If heated to above 135° C., the hysteresis loss soon attains a maximum, but then begins to decrease. Certain specimens heated to 160° C. were found to have their hysteresis loss doubled in a few days. The effect seems to come to a maximum at about 180° C. or 200° C. Mere lapse of time does not remove the increase, but if the iron is reannealed the augmentation in hysteresis disappears. If the iron is heated to a higher temperature, say between 300° C. and 700° C., Roget found the initial rise of hysteresis happens more quickly, but that the metal soon settles down into a state in which the hysteresis loss has a small but still augmented constant value. The augmentation in value, however, becomes more nearly zero as the temperature approaches 700° C. Brands of steel are now obtainable which do not age in this manner, but these *non-ageing* varieties of steel have not generally such low initial hysteresis values as the “Swedish Iron,” commonly considered best for the cores of transformers and alternating-current magnets.

The following conclusions have been reached in the matter:—(1) Iron and mild steel in the annealed state are more liable to change their hysteresis value by heating than when in the harder condition; (2) all changes are removed by re-annealing; (3) the changes thus produced by heating affect not only the amount of the hysteresis loss, but also the form of the lower part of the (B, H) curve.

Forms of Electromagnet.—The form which an electromagnet must take will greatly depend upon the purposes for which it is to be used. A design or form of electromagnet which will be very suitable for some purposes will be useless for others. Supposing it is desired to make an electromagnet which shall be capable of undergoing very rapid changes of strength, it must have such a form that the coercivity of the material is overcome by a self-demagnetizing force. This can be achieved by making the magnet in the form of a short and stout bar rather than a long thin one. It has already been explained that the ends or poles of a polar magnet exert a demagnetizing power upon the mass of the metal in the interior of the bar. If then the electromagnet has the form of a long thin bar, the length of which is several hundred times its diameter, the poles are very far removed from the centre of the bar, and the demagnetizing action will be very feeble; such a long thin electromagnet, although made of very soft iron, retains a considerable amount of magnetism after the magnetizing force is withdrawn. On the other hand, a very thick bar very quickly demagnetizes itself, because no part of the metal is far removed from the action of the free poles. Hence when, as in many telegraphic instruments, a piece of soft iron, called an armature, has to be attracted to the poles of a horseshoe-shaped electromagnet, this armature should be prevented from quite touching the polar surfaces of the magnet. If a soft iron mass does quite touch the poles, then it completes the magnetic circuit and abolishes the free poles, and the magnet is to a very large extent deprived of its self-demagnetizing power. This is the explanation of the well-known fact that after exciting the electromagnet and then stopping the current, it still requires a good pull to detach the “keeper”; but when once the keeper has been detached, the magnetism is found to have nearly disappeared. An excellent form of electromagnet for the production of very powerful fields has been designed by H. du Bois (fig. 6).

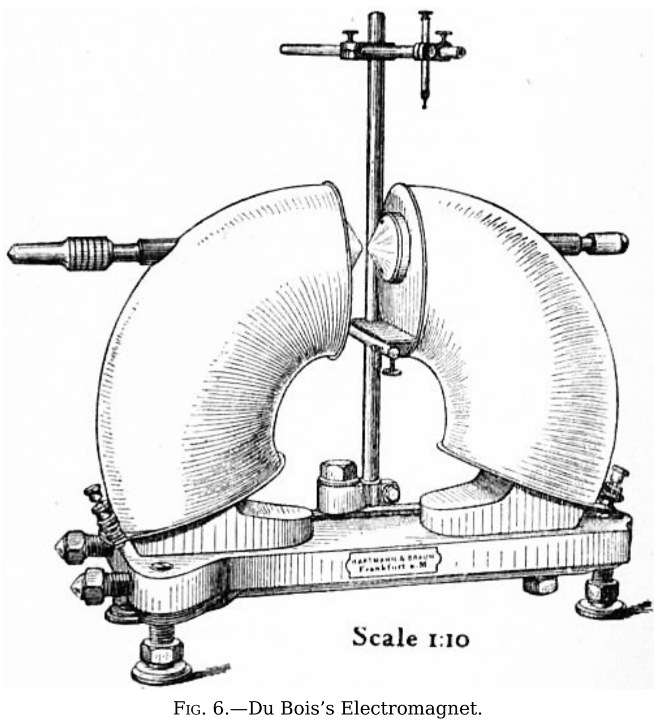

Fig. 6.—Du Bois’s Electromagnet.

Various forms of electromagnets used in connexion with dynamo machines are considered in the article Dynamo, and there is, therefore, no necessity to refer particularly to the numerous different shapes and types employed in electrotechnics.

Bibliography.—For additional information on the above subject the reader may be referred to the following works and original papers:—

H. du Bois, *The Magnetic Circuit in Theory and Practice*; S.P. Thompson, *The Electromagnet*; J.A. Fleming, *Magnets and Electric Currents*; J.A. Ewing, *Magnetic Induction in Iron and other Metals*; J.A. Fleming, "The Ferromagnetic Properties of Iron and Steel," *Proceedings of Sheffield Society of Engineers and Metallurgists* (Oct. 1897); J.A. Ewing, "The Magnetic Testing of Iron and Steel," *Proc. Inst. Civ. Eng.*, 1896, 126, p. 185; H.F. Parshall, "The Magnetic Data of Iron and Steel," *Proc. Inst. Civ. Eng.*, 1896, 126, p. 220; J.A. Ewing, "The Molecular Theory of Induced Magnetism," *Phil. Mag.*, Sept. 1890; W.M. Mordey, "Slow Changes in the Permeability of Iron," *Proc. Roy. Soc.* 57, p. 224; J.A. Ewing, "Magnetism," James Forrest Lecture, *Proc. Inst. Civ. Eng.* 138; S.P. Thompson, "Electromagnetic Mechanism," *Electrician*, 26, pp. 238, 269, 293; J.A. Ewing, "Experimental Researches in Magnetism," *Phil. Trans.*, 1885, part ii.; Ewing and Klassen, "Magnetic Qualities of Iron," *Proc. Roy. Soc.*, 1893.

(J. A. F.)

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1 In the *Annals of Philosophy* for November 1821 is a long article entitled "Electromagnetism" by Oersted, in which he gives a detailed account of his discovery. He had his thoughts turned to it as far back as 1813, but not until the 20th of July 1820 had he actually made his discovery. He seems to have been arranging a compass needle to observe any deflections during a storm, and placed near it a platinum wire through which a galvanic current was passed.

2 See *Trans. Soc. Arts*, 1825, 43, p. 38, in which a figure of Sturgeon's electromagnet is given as well as of other pieces of apparatus for which the Society granted him a premium and a silver medal.

3 See S.P. Thompson, *The Electromagnet* (London, 1891); J.A. Fleming, *A Handbook for the Electrical Laboratory and Testing Room*, vol. 2 (London, 1903); J.A. Ewing, *Magnetic Induction in Iron and other Metals* (London, 1903, 3rd ed.).

ELECTROMETALLURGY. The present article, as explained under Electrochemistry, treats only of those processes in which electricity is applied to the production of chemical reactions or molecular changes at furnace temperatures. In many of these the application of heat is necessary to bring the substances used into the liquid state for the purpose of electrolysis, aqueous solutions being unsuitable. Among the earliest experiments in this branch of the subject were those of Sir H. Davy, who in 1807 (*Phil. Trans.*, 1808, p. 1), produced the alkali metals by passing an intense current of electricity from a platinum wire to a platinum dish, through a mass of fused caustic alkali. The action was started in the cold, the alkali being slightly moistened to render it a conductor; then, as the current passed, heat was produced and the alkali fused, the metal being deposited in the liquid condition. Later, A. Matthiessen (*Quarterly Journ. Chem. Soc.* viii. 30) obtained potassium by the electrolysis of a mixture of potassium and calcium chlorides fused over a lamp. There are here foreshadowed two types of electrolytic furnace-operations: (*a*) those in which external heating maintains the electrolyte in the fused condition, and (*b*) those in which a current-density is applied sufficiently high to develop the heat necessary to effect this object unaided. Much of the earlier electro-metallurgical work was done with furnaces of the (*a*) type, while nearly all the later developments have been with those of class (*b*). There is a third class of operations, exemplified by the manufacture of calcium carbide, in which electricity is employed solely as a heating agent; these are termed *electrothermal*, as distinguished from *electrolytic*. In certain electrothermal processes (*e.g.* calcium carbide production) the heat from the current is employed in raising mixtures of substances to the temperature at which a desired chemical reaction will take place between them, while in others (*e.g.* the production of graphite from coke or gas-carbon) the heat is applied solely to the production of molecular or physical changes. In ordinary electrolytic work only the continuous current may of course be used, but in electrothermal work an alternating current is equally available.

Electric Furnaces.—Independently of the question of the application of external heating, the furnaces used in electrometallurgy may be broadly classified into (i.) arc furnaces, in which the intense heat of the electric arc is utilized, and (ii.) resistance and incandescence furnaces, in which the heat is generated by an electric current overcoming the resistance of an inferior conductor.

Arc furnaces.

Excepting such experimental arrangements as that of C.M. Despretz (*C.R.*, 1849, 29) for use on a small scale in the laboratory, Pichou in France and J.H. Johnson in England appear, in 1853, to have introduced the earliest practical form of furnace. In these arrangements, which were similar if not identical, the furnace charge was crushed to a fine powder and passed through two or more electric arcs in succession. When used for ore smelting, the reduced metal and the accompanying slag were to be caught, after leaving the arc and while still liquid, in a hearth fired with ordinary fuel. Although this primitive furnace could be made to act, its efficiency was low, and the use of a separate fire was disadvantageous. In 1878 Sir William Siemens patented a form of furnace[1] which is the type of a very large number of those designed by later inventors.

In the best-known form a plumbago crucible was used with a hole cut in the bottom to receive a carbon rod, which was ground in so as to make a tight joint. This rod was connected with the positive pole of the dynamo or electric generator. The crucible was fitted with a cover in which were two holes; one at the side to serve at once as sight-hole and charging door, the other in the centre to allow a second carbon rod to pass freely (without touching) into the interior. This rod was connected with the negative pole of the generator, and was suspended from one arm of a balance-beam, while from the other end of the beam was suspended a vertical hollow iron cylinder, which could be moved into or out of a wire coil or solenoid joined as a shunt across the two carbon rods of the furnace. The solenoid was above the iron cylinder, the supporting rod of which passed through it as a core. When the furnace with this well-known regulating device was to be used, say, for the melting of metals or other conductors of electricity, the fragments of metal were placed in the crucible and the positive electrode was brought near them. Immediately the current passed through the solenoid it caused the iron cylinder to rise, and, by means of its supporting rod, forced the end of the balance beam upwards, so depressing the other end that the negative carbon rod was forced downwards into contact with the metal in the crucible. This action completed the furnace-circuit, and current passed freely from the positive carbon through the fragments of metal to the negative carbon, thereby reducing the current through the shunt. At once the attractive force of the solenoid on the iron cylinder was automatically reduced, and the falling of the latter caused the negative carbon to rise, starting an arc between it and the metal in the crucible. A counterpoise was placed on the solenoid end of the balance beam to act against the attraction of the solenoid, the position of the counterpoise determining the length of the arc in the crucible. Any change in the resistance of the arc, either by lengthening, due to the sinking of the charge in the crucible, or by the burning of the carbon, affected the proportion of current flowing in the two shunt circuits, and so altered the position of the iron cylinder in the solenoid that the length of arc was, within limits, automatically regulated. Were it not for the use of some such device the arc would be liable to constant fluctuation and to frequent extinction. The crucible was surrounded with a bad conductor of heat to minimize loss by radiation. The positive carbon was in some cases replaced by a water-cooled metal tube, or ferrule, closed, of course, at the end inserted in the crucible. Several modifications were proposed, in one of which, intended for the heating of non-conducting substances, the electrodes were passed horizontally through perforations in the upper part of the crucible walls, and the charge in the lower part of the crucible was heated by radiation.

The furnace used by Henri Moissan in his experiments on reactions at high temperatures, on the fusion and volatilization of refractory materials, and on the formation of carbides, silicides and borides of various metals, consisted, in its simplest form, of two superposed blocks of lime or of limestone with a central cavity cut in the lower block, and with a corresponding but much shallower inverted cavity in the upper block, which thus formed the lid of the furnace. Horizontal channels were cut on opposite walls, through which the carbon poles or electrodes were passed into the upper part of the cavity. Such a furnace, to take a current of 4 H.P. (say, of 60 amperes and 50 volts), measured externally about 6 by 6 by 7 in., and the electrodes were about 0.4 in. in diameter, while for a current of 100 H.P. (say, of 746 amperes and 100 volts) it measured about 14 by 12 by 14 in., and the electrodes were about 1.5 in. in diameter. In the latter case the crucible, which was placed in the cavity immediately beneath the arc, was about 3 in. in diameter (internally), and about 3½ in. in height. The fact that energy is being used at so high a rate as 100 H.P. on so small a charge of material sufficiently indicates that the furnace is only used for experimental work, or for the fusion of metals which, like tungsten or chromium, can only be melted at temperatures attainable by electrical means. Moissan succeeded in fusing about ¾ ℔ of either of these metals in 5 or 6 minutes in a furnace similar to that last described. He also arranged an experimental tube-furnace by passing a carbon tube horizontally beneath the arc in the cavity of the lime blocks. When prolonged heating is required at very high temperatures it is 233
found necessary to line the furnace-cavity with alternate layers of magnesia and carbon, taking care that the lamina next to the lime is of magnesia; if this were not done the lime in contact with the carbon crucible would form calcium carbide and would slag down, but magnesia does not yield a carbide in this way. Chaplet has patented a muffle or tube furnace, similar in principle, for use on a larger scale, with a number of electrodes placed above and below the muffle-tube. The arc furnaces now widely used in the manufacture of calcium carbide on a large scale are chiefly developments of the Siemens furnace. But whereas, from its construction, the Siemens furnace was intermittent in operation, necessitating stoppage of the current while the contents of the crucible were poured out, many of the newer forms are specially designed either to minimize the time required in effecting the withdrawal of one charge and the introduction of the next, or to ensure absolute continuity of action, raw material being constantly charged in at the top and the finished substance and by-products (slag, &c.) withdrawn either continuously or at intervals, as sufficient quantity shall have accumulated. In the King furnace, for example, the crucible, or lowest part of the furnace, is made detachable, so that when full it may be removed and an empty crucible substituted. In the United States a revolving furnace is used which is quite continuous in action.

Incandescence furnaces.

The class of furnaces heated by electrically incandescent materials has been divided by Borchers into two groups: (1) those in which the substance is heated by contact with a substance offering a high resistance to the current passing through it, and (2) those in which the substance to be heated itself affords the resistance to the passage of the current whereby electric energy is converted into heat. Practically the first of these furnaces was that of Despretz, in which the mixture to be heated was placed in a carbon tube rendered incandescent by the passage of a current through its substance from end to end. In 1880 W. Borchers introduced his resistance-furnace, which, in one sense, is the converse of the Despretz apparatus. A thin carbon pencil, forming a bridge between two stout carbon rods, is set in the midst of the mixture to be heated. On passing a current through the carbon the small rod is heated to incandescence, and imparts heat to the surrounding mass. On a larger scale several pencils are used to make the connexions between carbon blocks which form the end walls of the furnace, while the side walls are of fire-brick laid upon one another without mortar. Many of the furnaces now in constant use depend mainly on this principle, a core of granular carbon fragments stamped together in the direct line between the electrodes, as in Acheson’s carborundum furnace, being substituted for the carbon pencils. In other cases carbon fragments are mixed throughout the charge, as in E.H. and A.H. Cowles’s zinc-smelting retort. In practice, in these furnaces, it is possible for small local arcs to be temporarily set up by the shifting of the charge, and these would contribute to the heating of the mass. In the remaining class of furnace, in which the electrical resistance of the charge itself is utilized, are the continuous-current furnaces, such as are used for the smelting of aluminium, and those alternating-current furnaces, (*e.g.* for the production of calcium carbide) in which a portion of the charge is first actually fused, and then maintained in the molten condition by the current passing through it, while the reaction between further portions of the charge is proceeding.

For ordinary metallurgical work the electric furnace, requiring as it does (excepting where waterfalls or other cheap sources of power are available) the intervention of the boiler and steam-engine, or of the gas or

Uses and advantages.

oil engine, with a consequent loss of energy, has not usually proved so economical as an ordinary direct fired furnace. But in some cases in which the current is used for electrolysis and for the production of extremely high temperatures, for which the calorific intensity of ordinary fuel is insufficient, the electric furnace is employed with advantage. The temperature of the electric furnace, whether of the arc or incandescence type, is practically limited to that at which the least easily vaporized material available for electrodes is converted into vapour. This material is carbon, and as its vaporizing point is (estimated at) over 3500° C., and less than 4000° C., the temperature of the electric furnace cannot rise much above 3500° C. (6330° F.); but H. Moissan showed that at this temperature the most stable of mineral combinations are dissociated, and the most refractory elements are converted into vapour, only certain borides, silicides and metallic carbides having been found to resist the action of the heat. It is not necessary that all electric furnaces shall be run at these high temperatures; obviously, those of the incandescence or resistance type may be worked at any convenient temperature below the maximum. The electric furnace has several advantages as compared with some of the ordinary types of furnace, arising from the fact that the heat is generated from within the mass of material operated upon, and (unlike the blast-furnace, which presents the same advantage) without a large volume of gaseous products of combustion and atmospheric nitrogen being passed through it. In ordinary reverberatory and other heating furnaces the burning fuel is without the mass, so that the vessel containing the charge, and other parts of the plant, are raised to a higher temperature than would otherwise be necessary, in order to compensate for losses by radiation, convection and conduction. This advantage is especially observed in some cases in which the charge of the furnace is liable to attack the containing vessel at high temperatures, as it is often possible to maintain the outer walls of the electric furnace relatively cool, and even to keep them lined with a protecting crust of unfused charge. Again, the construction of electric furnaces may often be exceedingly crude and simple; in the carborundum furnace, for example, the outer walls are of loosely piled bricks, and in one type of furnace the charge is simply heaped on the ground around the carbon resistance used for heating, without containing-walls of any kind. There is, however, one (not insuperable) drawback in the use of the electric furnace for the smelting of pure metals. Ordinarily carbon is used as the electrode material, but when carbon comes in contact at high temperatures with any metal that is capable of forming a carbide a certain amount of combination between them is inevitable, and the carbon thus introduced impairs the mechanical properties of the ultimate metallic product. Aluminium, iron, platinum and many other metals may thus take up so much carbon as to become brittle and unforgeable. It is for this reason that Siemens, Borchers and others substituted a hollow water-cooled metal block for the carbon cathode upon which the melted metal rests while in the furnace. Liquid metal coming in contact with such a surface forms a crust of solidified metal over it, and this crust thickens up to a certain point, namely, until the heat from within the furnace just overbalances that lost by conduction through the solidified crust and the cathode material to the flowing water. In such an arrangement, after the first instant, the melted metal in the furnace does not come in contact with the cathode material.

Electrothermal Processes.—In these processes the electric current is used solely to generate heat, either to induce chemical reactions between admixed substances, or to produce a physical (allotropic) modification of a given substance. Borchers predicted that, at the high temperatures available with the electric furnace, every oxide would prove to be reducible by the action of carbon, and this prediction has in most instances been justified. Alumina and lime, for example, which cannot be reduced at ordinary furnace temperatures, readily give up their oxygen to carbon in the electric furnace, and then combine with an excess of carbon to form metallic carbides. In 1885 the brothers Cowles patented a process for the electrothermal reduction of oxidized ores by exposure to an intense current of electricity when admixed with carbon in a retort. Later in that year they patented a process for the reduction of aluminium by carbon, and in 1886 an electric furnace with sliding carbon rods passed through the end walls to the centre of a rectangular furnace. The impossibility of working with just sufficient carbon to reduce the alumina, without using any excess which would be free to form at least so much carbide as would suffice, when diffused through the metal, to render it 234 brittle, practically restricts the use of such processes to the production of aluminium alloys.

Aluminium alloys.

Aluminium bronze (aluminium and copper) and ferro-aluminium (aluminium and iron) have been made in this way; the latter is the more satisfactory product, because a certain proportion of carbon is expected in an alloy of this character, as in ferromanganese and cast iron, and its presence is not objectionable. The furnace is built of fire-brick, and may measure (internally) 5 ft. in length by 1 ft. 8 in. in width, and 3 ft. in height. Into each end wall is built a short iron tube sloping downwards towards the centre, and through this is passed a bundle of five 3-in. carbon rods, bound together at the outer end by being cast into a head of cast iron for use with iron alloys, or of cast copper for aluminium bronze. This head slides freely in the cast iron tubes, and is connected by a copper rod with one of the terminals of the dynamo supplying the current. The carbons can thus, by the application of suitable mechanism, be withdrawn from or plunged into the furnace at will. In starting the furnace, the bottom is prepared by ramming it with charcoal-powder that has been soaked in milk of lime and dried, so that each particle is coated with a film of lime, which serves to reduce the loss of current by conduction through the lining when the furnace becomes hot. A sheet iron case is then placed within the furnace, and the space between it and the walls rammed with limed charcoal; the interior is filled with fragments of the iron or copper to be alloyed, mixed with alumina and coarse charcoal, broken pieces of carbon being placed in position to connect the electrodes. The iron case is then removed, the whole is covered with charcoal, and a cast iron cover with a central flue is placed above all. The current, either continuous or alternating, is then started, and continued for about 1 to 1½ hours, until the operation is complete, the carbon rods being gradually withdrawn as the action proceeds. In such a furnace a continuous current, for example, of 3000 amperes, at 50 to 60 volts, may be used at first, increasing to 5000 amperes in about half an hour. The reduction is not due to electrolysis, but to the action of carbon on alumina, a part of the carbon in the charge being consumed and evolved as carbon monoxide gas, which burns at the orifice in the cover so long as reduction is taking place. The reduced aluminium alloys itself immediately with the fused globules of metal in its midst, and as the charge becomes reduced the globules of alloy unite until, in the end, they are run out of the tap-hole after the current has been diverted to another furnace. It was found in practice (in 1889) that the expenditure of energy per pound of reduced aluminium was about 23 H.P.-hours, a number considerably in excess of that required at the present time for the production of pure aluminium by the electrolytic process described in the article Aluminium. Calcium carbide, graphite (*q.v.*), phosphorus (*q.v.*) and carborundum (*q.v.*) are now extensively manufactured by the operations outlined above.

Electrolytic Processes.—The isolation of the metals sodium and potassium by Sir Humphry Davy in 1807 by the electrolysis of the fused hydroxides was one of the earliest applications of the electric current to the extraction of metals. This pioneering work showed little development until about the middle of the 19th century. In 1852 magnesium was isolated electrolytically by R. Bunsen, and this process subsequently received much attention at the hands of Moissan and Borchers. Two years later Bunsen and H.E. Sainte Claire Deville working independently obtained aluminium (*q.v.*) by the electrolysis of the fused double sodium aluminium chloride. Since that date other processes have been devised and the electrolytic processes have entirely replaced the older methods of reduction with sodium. Methods have also been discovered for the electrolytic manufacture of calcium (*q.v.*), which have had the effect of converting a laboratory curiosity into a product of commercial importance. Barium and strontium have also been produced by electro-metallurgical methods, but the processes have only a laboratory interest at present. Lead, zinc and other metals have also been reduced in this manner.

For further information the following books, in addition to those mentioned at the end of the article Electrochemistry, may be consulted: Borchers, *Handbuch der Elektrochemie*; *Electric Furnaces* (Eng. trans. by H.G. Solomon, 1908); Moissan, *The Electric Furnace* (1904); J. Escard, *Fours électriques* (1905); *Les Industries électrochimiques* (1907).

(W. G. M.)

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1 Cf. Siemens's account of the use of this furnace for experimental purposes in *British Association Report* for 1882.

ELECTROMETER, an instrument for measuring difference of potential, which operates by means of electrostatic force and gives the measurement either in arbitrary or in absolute units (see Units, Physical). In the last case the instrument is called an absolute electrometer. Lord Kelvin has classified electrometers into (1) Repulsion, (2) Attracted disk, and (3) Symmetrical electrometers (see W. Thomson, *Brit. Assoc. Report,* 1867, or *Reprinted Papers on Electrostatics and Magnetization*, p. 261).

Repulsion Electrometers.—The simplest form of repulsion electrometer is W. Henley's pith ball electrometer (*Phil. Trans.*, 1772, 63, p. 359) in which the repulsion of a straw ending in a pith ball from a fixed stem is indicated on a graduated arc (see Electroscope). A double pith ball repulsion electrometer was employed by T. Cavallo in 1777.

It may be pointed out that such an arrangement is not merely an arbitrary electrometer, but may become an absolute electrometer within certain rough limits. Let two spherical pith balls of radius r and weight W, covered with gold-leaf so as to be conducting, be suspended by parallel silk threads of length l so as just to touch each other. If then the balls are both charged to a potential V they will repel each other, and the threads will stand out at an angle 2θ, which can be observed on a protractor. Since the electrical repulsion of the balls is equal to $C^2V^2 4l^2 \sin^2\theta$ dynes, where C = r is the capacity of either ball, and this force is balanced by the restoring force due to their weight, Wg dynes, where g is the acceleration of gravity, it is easy to show that we have

$$V = \frac{2l \sin\theta \sqrt{Wg \tan\theta}}{r}$$

as an expression for their common potential V, provided that the balls are small and their distance sufficiently great not sensibly to disturb the uniformity of electric charge upon them. Observation of θ with measurement of the value of l and r reckoned in centimetres and W in grammes gives us the potential difference of the balls in absolute C.G.S. or electrostatic units. The gold-leaf electroscope invented by Abraham Bennet (see Electroscope) can in like manner, by the addition of a scale to observe the divergence of the gold-leaves, be made a repulsion electrometer.

Attracted Disk Electrometers.—A form of attracted disk absolute electrometer was devised by A. Volta. It consisted of a plane conducting plate forming one pan of a balance which was suspended over another insulated plate which could be electrified. The attraction between the two plates was balanced by a weight put in the opposite pan. A similar electric balance was subsequently devised by Sir W. Snow-Harris,[1] one of whose instruments is shown in fig. 1. C is an insulated disk over which is suspended another disk attached to the arm of a balance. A weight is put in the opposite scale pan and a measured charge of electricity is given to the disk C just sufficient to tip over the balance. Snow-Harris found that this charge varied as the square root of the weight in the opposite pan, thus showing that the attraction between the disks at given distance apart varies as the square of their difference of potential.

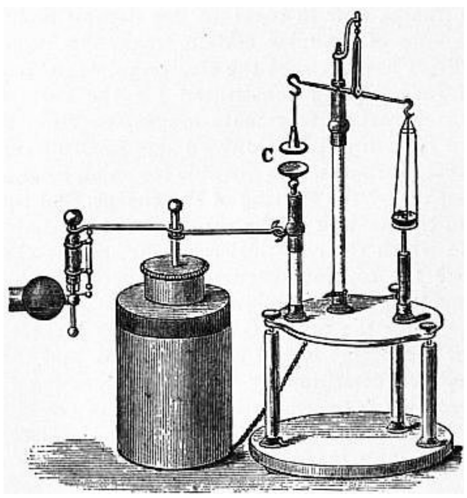


Fig. 1.—Snow-Harris's Disk Electrometer.

The most important improvements in connexion with electrometers are due, however, to Lord Kelvin, who introduced the guard plate and used gravity or the torsion of a wire as a means for evaluating the electrical forces.

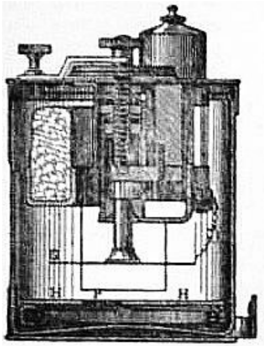


FIG. 2.—Kelvin's Portable Electrometer.

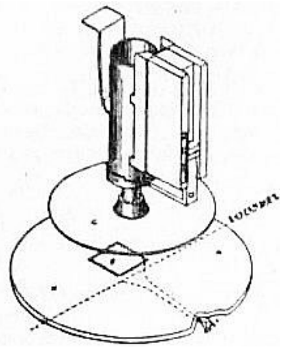


FIG. 3.

His portable electrometer is shown in fig. 2. H H (see fig. 3) is a plane disk of metal called the guard plate, fixed to the inner coating of a small Leyden jar (see fig. 2). At F a square hole is cut out of H H, and into this fits loosely without touching, like a trap door, a square piece of aluminium foil having a projecting tail, which carries at its end a stirrup L, crossed by a fine hair (see fig. 3). The square piece of aluminium is pivoted round a horizontal stretched wire. If then another horizontal disk G is placed over the disk H H and a difference of potential made between G and H H, the movable aluminium trap door F will be attracted by the fixed plate G. Matters are so arranged by giving a torsion to the wire carrying the aluminium disk F that for a certain potential difference between the plates H and G, the movable part F comes into a definite sighted position, which is observed by means of a small lens. The plate G (see fig. 2) is moved up and down, parallel to itself, by means of a screw. In using the instrument the conductor, whose potential is to be tested, is connected to the plate G. Let this potential be denoted by V, and let v be the potential of the guard plate and the aluminium flap. This last potential is maintained constant by guard plate and flap being part of the interior coating of a charged Leyden jar. Since the distribution of electricity may be considered to be constant over the surface S of the attracted disk, the mechanical force f on it is given by the expression,[2]

$$f = \frac{S\,(V - v)^2}{8\pi d^2},$$

where d is the distance between the two plates. If this distance is varied until the attracted disk comes into a definite sighted position as seen by observing the end of the index through the lens, then since the force f is constant, being due to the torque applied by the wire for a definite angle of twist, it follows that the difference of potential of the two plates varies as their distance. If then two experiments are made, first with the upper plate connected to earth, and secondly, connected to the object being tested, we get an expression for the potential V of this conductor in the form

$$V = A\,(d' - d),$$

where d and d′ are the distances of the fixed and movable plates from one another in the two cases, and A is some constant. We thus find V in terms of the constant and the difference of the two screw readings.

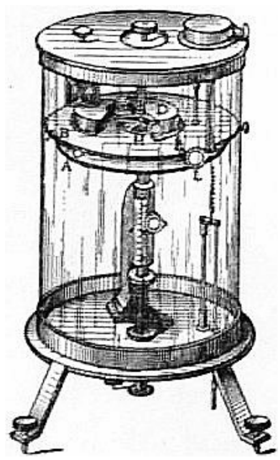


FIG. 4.—Kelvin's Absolute Electrometer.

Lord Kelvin's absolute electrometer (fig. 4) involves the same principle. There is a certain fixed guard disk B having a hole in it which is loosely occupied by an aluminium trap door plate, shielded by D and suspended on springs, so that its surface is parallel with that of the guard plate. Parallel to this is a second movable plate A, the distances between the two being measurable by means of a screw. The movable plate can be drawn down into a definite sighted position when a difference of potential is made between the two plates. This sighted position is such that the surface of the trap door plate is level with that of the guard plate, and is determined by observations made with the lenses H and L. The movable plate can be thus depressed by placing on it a certain standard weight W grammes.

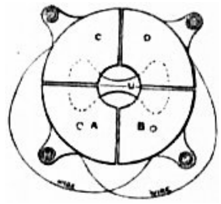
Suppose it is required to measure the difference of potentials V and V′ of two conductors. First one and then the other conductor is connected with the electrode of the lower or movable plate, which is moved by the screw until the index attached to the attracted disk shows it to be in the sighted position. Let the screw readings in the two cases be d and d′. If W is the weight required to depress the attracted disk into the same sighted position when the plates are unelectrified and g is the acceleration of gravity, then the difference of potentials of the conductors tested is expressed by the formula

$$V - V' = (d - d')\sqrt{\frac{8\pi g W}{S}},$$

where S denotes the area of the attracted disk.

The difference of potentials is thus determined in terms of a weight, an area and a distance, in absolute C.G.S. measure or electrostatic units.

Symmetrical Electrometers include the dry pile electrometer and Kelvin's quadrant electrometer. The principle underlying these instruments is that we can measure differences of potential by means of the motion of an electrified body in a symmetrical field of electric force. In the dry pile electrometer a single gold-leaf is hung up between two plates which are connected to the opposite terminals of a dry pile so that a certain constant difference of potential exists between these plates. The original inventor of this instrument was T.G.B. Behrens (*Gilb. Ann.*, 1806, 23), but it generally bears the name of J.G.F. von Bohnenberger, who slightly modified



its form. G.T. Fechner introduced the important improvement of using only one pile, which he removed from the immediate neighbourhood of the suspended leaf. W.G. Hankel still further improved the dry pile electrometer by giving a slow motion movement to the two plates, and substituted a galvanic battery with a large number of cells for the dry pile, and also employed a divided scale to measure the movements of the gold-leaf (*Pogg. Ann.*, 1858, 103). If the gold-leaf is unelectrified, it is not acted upon by the two plates placed at equal distances on either side of it, but if its potential is raised or lowered it is attracted by one disk and repelled by the other, and the displacement becomes a measure of its potential.

FIG. 5.

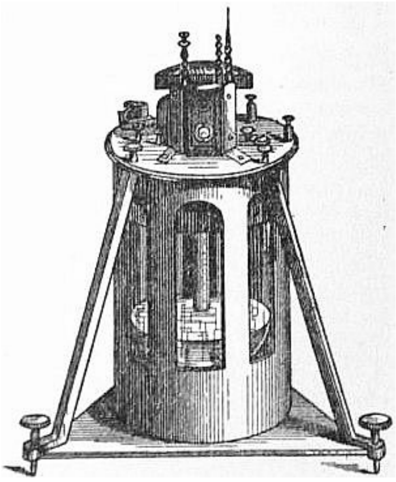


FIG. 6.—Kelvin's Quadrant Electrometer.

A vast improvement in this instrument was made by the invention of the quadrant electrometer by Lord Kelvin, which is the most sensitive form of electrometer yet devised. In this instrument (see fig. 5) a flat paddle-shaped needle of aluminium foil U is supported by a bifilar suspension consisting of two cocoon fibres. This needle is suspended in the interior of a glass vessel partly coated with tin-foil on the outside and inside, forming therefore a Leyden jar (see fig. 6). In the bottom of the vessel is placed some sulphuric acid, and a platinum wire attached to the suspended needle dips into this acid. By giving a charge to this Leyden jar the needle can thus be maintained at a certain constant high potential. The needle is enclosed by a sort of flat box divided into four insulated quadrants A, B, C, D (fig. 5), whence the name. The opposite quadrants are connected together by thin platinum wires. These quadrants are insulated from the 236
needle and from the case, and the two pairs are connected to two electrodes. When the instrument is to be used to determine the potential difference between two conductors, they are connected to the two opposite pairs of quadrants. The needle in its normal position is symmetrically placed with regard to the quadrants, and carries a mirror by means of which its displacement can be observed in the usual manner by reflecting the ray of light from it. If the two quadrants are at different potentials, the needle moves from one quadrant towards the other, and the image of a spot of light on the scale is therefore displaced. Lord Kelvin provided the instrument with two necessary adjuncts, viz. a replenisher or rotating electrophorus (*q.v.*), by means of which the charge of the Leyden jar which forms the enclosing vessel can be increased or diminished, and also a small aluminium balance plate or gauge, which is in principle the same as the attracted disk portable electrometer by means of which the potential of the inner coating of the Leyden jar is preserved at a known value.

According to the mathematical theory of the instrument,[3] if V and V′ are the potentials of the quadrants and v is the potential of the needle, then the torque acting upon the needle to cause rotation is given by the expression,

$$C (V - V') \{v - \tfrac{1}{2} (V + V')\},$$

where C is some constant. If v is very large compared with the mean value of the potentials of the two quadrants, as it usually is, then the above expression indicates that the couple varies as the difference of the potentials between the quadrants.

Dr J. Hopkinson found, however, before 1885, that the above formula does not agree with observed facts (*Proc. Phys. Soc. Lond.*, 1885, 7, p. 7). The formula indicates that the sensibility of the instrument should increase with the charge of the Leyden jar or needle, whereas Hopkinson found that as the potential of the needle was increased by working the replenisher of the jar, the deflection due to three volts difference between the quadrants first increased and then diminished. He found that when the potential of the needle exceeded a certain value, of about 200 volts, for the particular instrument he was using (made by White of Glasgow), the above formula did not hold good. W.E. Ayrton, J. Perry and W.E. Sumpner, who in 1886 had noticed the same fact as Hopkinson, investigated the matter in 1891 (*Proc. Roy. Soc.*, 1891, 50, p. 52; *Phil. Trans.*, 1891, 182, p. 519). Hopkinson had been inclined to attribute the anomaly to an increase in the tension of the bifilar threads, owing to a downward pull on the needle, but they showed that this theory would not account for the discrepancy. They found from observations that the particular quadrant electrometer they used might be made to follow one or other of three distinct laws. If the quadrants were near together there were certain limits between which the potential of the needle might vary without producing more than a small change in the deflection corresponding with the fixed potential difference of the quadrants. For example, when the quadrants were about 2.5 mm. apart and the suspended fibres near together at the top, the deflection produced by a P.D. of 1.45 volts between the quadrants only varied about 11% when the potential of the needle varied from 896 to 3586 volts. When the fibres were far apart at the top a similar flatness was obtained in the curve with the quadrants about 1 mm. apart. In this case the deflection of the needle was practically quite constant when its potential varied from 2152 to 3227 volts. When the quadrants were about 3.9 mm. apart, the deflection for a given P.D. between the quadrants was almost directly proportional to the potential of the needle. In other words, the electrometer nearly obeyed the theoretical law. Lastly, when the quadrants were 4 mm. or more apart, the deflection increased much more rapidly than the potential, so that a maximum sensibility bordering on instability was obtained. Finally, these observers traced the variation to the fact that the wire supporting the aluminium needle as well as the wire which connects the needle with the sulphuric acid in the Leyden jar in the White pattern of Leyden jar is enclosed in a metallic guard tube to screen the wire from external action. In order that the needle may project outside the guard tube, openings are made in its two sides; hence the moment the needle is deflected each half of it becomes unsymmetrically placed relatively to the two metallic pieces which join the upper and lower half of the guard tube. Guided by these experiments, Ayrton, Perry and Sumpner constructed an improved unifilar quadrant electrometer which was not only more sensitive than the White pattern, but fulfilled the theoretical law of working. The bifilar suspension was abandoned, and instead a new form of adjustable magnetic control was adopted. All the working parts of the instrument were supported on the base, so that on removing a glass shade which serves as a Leyden jar they can be got at and adjusted in position. The conclusion to which the above observers came was that any quadrant electrometer made in any manner does not necessarily obey a law of deflection making the deflections proportional to the potential difference of the quadrants, but that an electrometer can be constructed which does fulfil the above law.

The importance of this investigation resides in the fact that an electrometer of the above pattern can be used as a wattmeter (*q.v.*), provided that the deflection of the needle is proportional to the potential difference of the quadrants. This use of the instrument was proposed simultaneously in 1881 by Professors Ayrton and G.F. Fitzgerald and M.A. Potier. Suppose we have an inductive and a non-inductive circuit in series, which is traversed by a periodic current, and that we desire to know the power being absorbed to the inductive circuit. Let v_1, v_2, v_3 be the instantaneous potentials of the two ends and middle of the circuit; let a quadrant electrometer be connected first with the quadrants to the two ends of the inductive circuit and the needle to the far end of the non-inductive circuit, and then secondly with the needle connected to one of the quadrants (see fig. 5). Assuming the electrometer to obey the above-mentioned theoretical law, the first reading is proportional to

$$v_1 - v_2 \left\{ v_3 - \frac{v_1 + v_2}{2} \right\}$$

and the second to

$$v_1 - v_2 \left\{ v_2 - \frac{v_1 + v_2}{2} \right\}.$$

The difference of the readings is then proportional to

$$(v_1 - v_2)(v_2 - v_3).$$

But this last expression is proportional to the instantaneous power taken up in the inductive circuit, and hence the difference of the two readings of the electrometer is proportional to the mean power taken up in the circuit (*Phil. Mag.*, 1891, 32, p. 206). Ayrton and Perry and also P.R. Blondlot and P. Curie afterwards suggested that a single electrometer could be constructed with two pairs of quadrants and a duplicate needle on one stem, so as to make two readings simultaneously and produce a deflection proportional at once to the power being taken up in the inductive circuit.

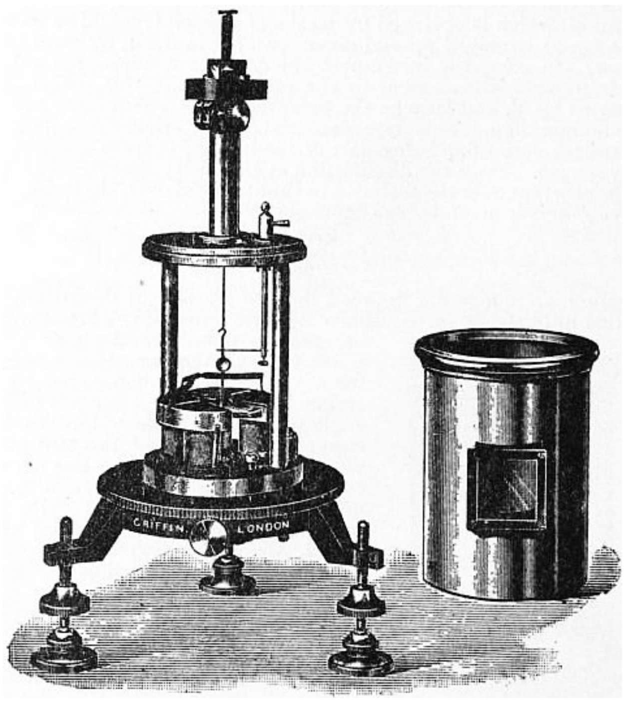


Fig. 7.—Quadrant Electrometer. Dolezalek Pattern.

Quadrant electrometers have also been designed especially for measuring extremely small potential differences. An instrument of this kind has been constructed by Dr. F. Dolezalek (fig. 7). The needle and quadrants are of small size, and the electrostatic capacity is correspondingly small. The quadrants are mounted on pillars of amber which afford a very high insulation. The needle, a piece of paddle-shaped paper thinly coated with silver foil, is suspended by a quartz fibre, its extreme lightness making it possible to use a very feeble controlling force without rendering the period of oscillation unduly great. The resistance offered by the air to a needle of such light construction suffices to render the motion nearly dead-beat. Throughout a wide range the deflections are proportional to the potential difference producing them. The needle is charged to a potential of 50 to 200 volts by means of a dry pile or voltaic battery, or from a lighting circuit. 237
To facilitate the communication of the charge to the needle, the quartz fibre and its attachments are rendered conductive by a thin film of solution of hygroscopic salt such as calcium chloride. The lightness of the needle enables the instrument to be moved without fear of damaging the suspension. The upper end of the quartz fibre is rotated by a torsion head, and a metal cover serves to screen the instrument from stray electrostatic fields. With a quartz fibre 0.009 mm. thick and 60 mm. long, the needle being charged to 110 volts, the period and swing of the needle was 18 seconds. With the scale at a distance of two metres, a deflection of 130 mm. was produced by an electromotive force of 0.1 volt. By using a quartz fibre of about half the above diameter the sensitiveness was much increased. An instrument of this form is valuable in measuring small alternating currents by the fall of potential produced down a known resistance. In the same

way it may be employed to measure high potentials by measuring the fall of potential down a fraction of a known non-inductive resistance. In this last case, however, the capacity of the electrometer used must be small, otherwise an error is introduced.[4]

See, in addition to references already given, A. Gray, *Absolute Measurements in Electricity and Magnetism* (London, 1888), vol. i. p. 254; A. Winkelmann, *Handbuch der Physik* (Breslau, 1905), pp. 58-70, which contains a large number of references to original papers on electrometers.

(J. A. F.)

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1 It is probable that an experiment of this kind had been made as far back as 1746 by Daniel Gralath, of Danzig, who has some claims to have suggested the word “electrometer” in connexion with it. See Park Benjamin, *The Intellectual Rise in Electricity* (London, 1895), p. 542.

2 See Maxwell, *Treatise on Electricity and Magnetism* (2nd ed.), i. 308.

3 See Maxwell, *Electricity and Magnetism* (2nd ed., Oxford, 1881), vol. i. p. 311.

4 See J.A. Fleming, *Handbook for the Electrical Laboratory and Testing Room*, vol. i. p. 448 (London, 1901).

ELECTRON, the name suggested by Dr G. Johnstone Stoney in 1891 for the natural unit of electricity to which he had drawn attention in 1874, and subsequently applied to the ultra-atomic particles carrying negative charges of electricity, of which Professor Sir J.J. Thomson proved in 1897 that the cathode rays consisted. The electrons, which Thomson at first called corpuscles, are point charges of negative electricity, their inertia showing them to have a mass equal to about $1/2000$ that of the hydrogen atom. They are apparently derivable from all kinds of matter, and are believed to be components at any rate of the chemical atom. The electronic theory of the chemical atom supposes, in fact, that atoms are congeries of electrons in rapid orbital motion. The size of the electron is to that of an atom roughly in the ratio of a pin’s head to the dome of St Paul’s cathedral. The electron is always associated with the unit charge of negative electricity, and it has been suggested that its inertia is wholly electrical. For further details see the articles on Electricity; Magnetism; Matter; Radioactivity; Conduction, Electric; *The Electron Theory*, E. Fournier d’Albe (London, 1907); and the original papers of Dr G. Johnstone Stoney, *Proc. Brit. Ass.* (Belfast, August 1874), “On the Physical Units of Nature,” and *Trans. Royal Dublin Society* (1891), 4, p. 583.

ELECTROPHORUS, an instrument invented by Alessandro Volta in 1775, by which mechanical work is transformed into electrostatic charge by the aid of a small initial charge of electricity. The operation depends on the facts of electrostatic induction discovered by John Canton in 1753, and, independently, by J.K. Wilcke in 1762 (see Electricity). Volta, in a letter to J. Priestley on the 10th of June 1775 (see *Collezione dell’ opere*, ed. 1816, vol. i. p. 118), described the invention of a device he called an *elettroforo perpetuo*, based on the fact that a conductor held near an electrified body and touched by the finger was found, when withdrawn, to possess an electric charge of opposite sign to that of the electrified body. His electrophorus in one form consisted of a disk of non-conducting material, such as pitch or resin, placed between two metal sheets, one being provided with an insulating handle. For the pitch or resin may be substituted a sheet of glass, ebonite, india-rubber or any other good dielectric placed upon a metallic sheet, called the sole-plate. To use the apparatus the surface of the dielectric is rubbed with a piece of warm flannel, silk or catskin, so as to electrify it, and the upper metal plate is then placed upon it. Owing to the irregularities in the surfaces of the dielectric and upper plate the two are only in contact at a few points, and owing to the insulating quality of the dielectric its surface electrical charge cannot move over it. It therefore acts inductively upon the upper plate and induces on the adjacent surface an electric charge of opposite sign. Suppose, for instance, that the dielectric is a plate of resin rubbed with catskin, it will then be negatively electrified and will act by induction on the upper plate across the film of air separating the upper resin surface and lower surface of the upper metal plate. If the upper plate is touched with the finger or connected to earth for a moment, a negative charge will escape from the metal plate to earth at that moment. The arrangement thus constitutes a condenser; the upper plate on its under surface carries a charge of positive electricity and the resin plate a charge of negative electricity on its upper surface, the air film between them being the dielectric of the condenser. If, therefore, the upper plate is elevated, mechanical work has to be done to separate the two electric charges. Accordingly on raising the upper plate, the charge on it, in old-fashioned nomenclature, becomes *free* and can be communicated to any other insulated conductor at a lower potential, the upper plate thereby becoming more or less discharged. On placing the upper plate again on the resin and touching it for a moment, the process can be repeated, and so at the expense of mechanical work done in lifting the upper plate against the mutual attraction of two electric charges of opposite sign, an indefinitely large electric charge can be accumulated and given to any other suitable conductor. In course of time, however, the surface charge of the resin becomes dissipated and it then has to be again excited. To avoid the necessity for touching the upper plate every time it is put down on the resin, a metal pin may be brought through the insulator from the sole-plate so that each time that the upper plate is put down on the resin it is automatically connected to earth. We are thus able by a process of merely lifting the upper plate repeatedly to convey a large electrical charge to some conductor starting from the small charge produced by friction on the resin. The above explanation does not take into account the function of the sole-plate, which is important. The sole-plate serves to increase the electrical capacity of the upper plate when placed down upon the resin or excited insulator. Hence when so placed it takes a larger charge. When touched by the finger the upper plate is brought to zero potential. If then the upper plate is lifted by its insulating handle its capacity becomes diminished. Since, however, it carries with it the charge it had when resting on the resin, its potential becomes increased as its capacity becomes less, and it therefore rises to a high potential, and will give a spark if the knuckle is approached to it when it is lifted after having been touched and raised.

The study of Volta's electrophorus at once suggested the performance of these cyclical operations by some form of rotation instead of elevation, and led to the invention of various forms of doubler or multiplier. The instrument was thus the first of a long series of machines for converting mechanical work into electrostatic energy, and the predecessor of the modern type of influence machine (see Electrical Machine). Volta himself devised a double and reciprocal electrophorus and also made mention of the subject of multiplying condensers in a paper published in the *Phil. Trans.* for 1782 (p. 237, and appendix, p. vii.). He states, however, that the use of a condenser in connexion with an electrophorus to make evident and multiply weak charges was due to T. Cavallo (*Phil. Trans.*, 1788).

For further information see S.P. Thompson, "The Influence Machine from 1788 to 1888," *Journ. Inst. Tel. Eng.*, 1888, 17, p. 569. Many references to original papers connected with the electrophorus will be found in A. Winkelmann's *Handbuch der Physik* (Breslau, 1905), vol. iv. p. 48.

(J. A. F.)

ELECTROPLATING, the art of depositing metals by the electric current. In the article Electrolysis it is shown how the passage of an electric current through a solution containing metallic ions involves the deposition of the metal on the cathode. Sometimes the metal is deposited in a pulverulent form, at others as a firm tenacious film, the nature of the deposit being dependent upon the particular metal, the concentration of the solution, the difference of potential between the electrodes, and other experimental conditions. As the durability of the electro-deposited coat on plated wares of all kinds is of the utmost importance, the greatest 238 care must be taken to ensure its complete adhesion. This can only be effected if the surface of the metal on which the deposit is to be made is chemically clean. Grease must be removed by potash, whiting or other means, and tarnish by an acid or potassium cyanide, washing in plenty of water being resorted to after each operation. The vats for depositing may be of enamelled iron, slate, glazed earthenware, glass, lead-lined wood, &c. The current densities and potential differences frequently used for some of the commoner metals are given in the following table, taken from M'Millan's *Treatise on Electrometallurgy*. It must be remembered, however, that variations in conditions modify the electromotive force required for any given process. For example, a rise in temperature of the bath causes an increase in its conductivity, so that a lower E.M.F. will suffice to give the required current density; on the other hand, an abnormally great distance between the electrodes, or a diminution in acidity of an acid bath, or in the strength of the solution used, will increase the resistance, and so require the application of a higher E.M.F.

Metal.	Amperes.		Volts between Anode and Cathode.
	Per sq. decimetre of Cathode Surface.	Per sq. in. of Cathode Surface.	
Antimony	0.4-0.5	0.02-0.03	1.0-1.2
Brass	0.5-0.8	0.03-0.05	3.0-4.0
Copper, acid bath	1.0-1.5	0.065-0.10	0.5-1.5
Copper, alkaline bath	0.3-0.5	0.02-0.03	3.0-5.0
Gold	0.1	0.006	0.5-4.0
Iron	0.5	0.03	1.0
Nickel, at first	1.4-1.5	0.09-0.10	5.0
Nickel, after	0.2-0.3	0.015-0.02	1.5-2.0
Nickel, on zinc	0.4	0.025	4.0-5.0
Silver	0.2-0.5	0.015-0.03	0.75-1.0
Zinc	0.3-0.6	0.02-0.04	2.5-3.0

Large objects are suspended in the tanks by hooks or wires, care being taken to shift their position and so avoid wire-marks. Small objects are often heaped together in perforated trays or ladles, the cathode connecting-rod being buried in the midst of them. These require constant shifting because the objects are in contact at many points, and because the top ones shield those below from the depositing action of the current. Hence processes have been patented in which the objects to be plated are suspended in revolving drums between the anodes, the rotation of the drum causing the constant renewal of surfaces and affording a burnishing action at the same time. Care must be taken not to expose goods in the plating-bath to too high a current density, else they may be "burnt"; they must never be exposed one at a time to the full anode surface, with the current flowing in an empty bath, but either one piece at a time should be replaced, or some of the anodes should be transferred temporarily to the place of the cathodes, in order to distribute the current over a sufficient cathode-area. Burnt deposits are dark-coloured, or even pulverulent and useless. The strength of the current may also be regulated by introducing lengths of German silver or iron wire, carbon rod, or other inferior conductors in the path of the current, and a series of such resistances should always be provided close to the tanks. Ammeters to measure the volume, and voltmeters to determine the pressure of current supplied to the baths, should also be provided. Very irregular surfaces may require the use of specially shaped anodes in order that the distance between the electrodes may be fairly uniform, otherwise the portion of the cathode lying nearest to the anode may receive an undue share of the current, and therefore a greater thickness of coat. Supplementary anodes are sometimes used in difficult cases of this kind. Large metallic surfaces (especially external surfaces) are sometimes plated by means of a “doctor,” which, in its simplest form, is a brush constantly wetted with the electrolyte, with a wire anode buried amid the hairs or bristles; this brush is painted slowly over the surface of the metal to be coated, which must be connected to the negative terminal of the electrical generator. Under these conditions electrolysis of the solution in the brush takes place. Iron ships’ plates have recently been coated with copper in sections (to prevent the adhesion of barnacles), by building up a temporary trough against the side of the ship, making the thoroughly cleansed plate act both as cathode and as one side of the trough. Decorative plating-work in several colours (*e.g.* “parcel-gilding”) is effected by painting a portion of an object with a stopping-out (*i.e.* a non-conducting) varnish, such as copal varnish, so that this portion is not coated. The varnish is then removed, a different design stopped out, and another metal deposited. By varying this process, designs in metals of different colours may readily be obtained.

Reference must be made to the textbooks (see Electrochemistry) for a fuller account of the very varied solutions and methods employed for electroplating with silver, gold, copper, iron and nickel. It should be mentioned here, however, that solutions which would deposit their metal on any object by simple immersion should not be generally used for electroplating that object, as the resulting deposit is usually non-adhesive. For this reason the acid copper-bath is not used for iron or zinc objects, a bath containing copper cyanide or oxide dissolved in potassium cyanide being substituted. This solution, being an inferior conductor of electricity, requires a much higher electromotive force to drive the current through it, and is therefore more costly in use. It is, however, commonly employed hot, whereby its resistance is reduced. *Zinc* is commonly deposited by electrolysis on iron or steel goods which would ordinarily be “galvanized,” but which for any reason may not conveniently be treated by the method of immersion in fused zinc. The zinc cyanide bath may be used for small objects, but for heavy goods the sulphate bath is employed. Sherard Cowper-Coles patented a process in which, working with a high current density, a lead anode is used, and powdered zinc is kept suspended in the solution to maintain the proportion of zinc in the electrolyte, and so to guard against the gradual acidification of the bath. *Cobalt* is deposited by a method analogous to that used for its sister-metal nickel. *Platinum*, *palladium* and *tin* are occasionally deposited for special purposes. In the deposition of *gold* the colour of the deposit is influenced by the presence of impurities in the solution; when copper is present, some is deposited with the gold, imparting to it a reddish colour, whilst a little silver gives it a greenish shade. Thus so-called coloured-gold deposits may be produced by the judicious introduction of suitable impurities. Even pure gold, it may be noted, is darker or lighter in colour according as a stronger or a weaker current is used. The electro-deposition of *brass*—mainly on iron ware, such as bedstead tubes—is now very widely practised, the bath employed being a mixture of copper, zinc and potassium cyanides, the proportions of which vary according to the character of the brass required, and to the mode of treatment. The colour depends in part upon the proportion of copper and zinc, and in part upon the current density, weaker currents tending to produce a redder or yellower metal. Other alloys may be produced, such as bronze, or German silver, by selecting solutions (usually cyanides) from which the current is able to deposit the constituent metals simultaneously.

Electrolysis has in a few instances been applied to processes of manufacture. For example, Wilde produced copper printing surfaces for calico printing-rollers and the like by immersing rotating iron cylinders as cathodes in a copper bath. Elmore, Dumoulin, Cowper-Coles and others have prepared copper cylinders and plates by depositing copper on rotating mandrels with special arrangements. Others have arranged a means of obtaining high conductivity wire from cathode-copper without fusion, by depositing the metal in the form of a spiral strip on a cylinder, the strip being subsequently drawn down in the usual way; at present, however, the ordinary methods of wire production are found to be cheaper. J.W. Swan (*Journ. Inst. Elec.* 239 *Eng.*, 1898, vol. xxvii. p. 16) also worked out, but did not proceed with, a process in which a copper wire whilst receiving a deposit of copper was continuously passed through the draw-plate, and thus indefinitely extended in length. Cowper-Coles (*Journ. Inst. Elec. Eng.*, 1898, 27, p. 99) very successfully produced true parabolic reflectors for projectors, by depositing copper upon carefully ground and polished glass surfaces rendered conductive by a film of deposited silver.

ELECTROSCOPE, an instrument for detecting differences of electric potential and hence electrification. The earliest form of scientific electroscope was the *versorium* or electrical needle of William Gilbert (1544-1603), the celebrated author of the treatise *De magnete* (see Electricity). It consisted simply of a light metallic needle balanced on a pivot like a compass needle. Gilbert employed it to prove that numerous other bodies besides amber are susceptible of being electrified by friction.[1] In this case the visible indication consisted in the attraction exerted between the electrified body and the light pivoted needle which was acted upon and electrified by induction. The next improvement was the invention of simple forms of repulsion electroscope. Two similarly electrified bodies repel each other. Benjamin Franklin employed the repulsion of two linen threads, C.F. de C. du Fay, J. Canton, W. Henley and others devised the pith ball, or double straw electroscope (fig. 1). T. Cavallo about 1770 employed two fine silver wires terminating in pith balls suspended in a glass vessel having strips of tin-foil pasted down the sides (fig. 2). The object of the thimble-shaped dome was to keep moisture from the stem from which the pith balls were supported, so that the apparatus could be used in the open air even in the rainy weather. Abraham Bennet (*Phil. Trans.*, 1787, 77, p. 26) invented the modern form of gold-leaf electroscope. Inside a glass shade he fixed to an insulated wire a pair of strips of gold-leaf (fig. 3). The wire terminated in a plate or knob outside the vessel. When an electrified body was held near or in contact with the knob, repulsion of the gold leaves ensued. Volta added the condenser (*Phil. Trans.*, 1782), which greatly increased the power of the instrument. M. Faraday, however, showed long subsequently that to bestow upon the indications of such an electroscope definite meaning it was necessary to place a cylinder of metallic gauze connected to the earth inside the vessel, or better still, to line the glass shade with tin-foil connected to the earth and observe through a hole the indications of the gold leaves (fig. 4). Leaves of aluminium foil may with advantage be substituted for gold-leaf, and a scale is sometimes added to indicate the angular divergence of the leaves.

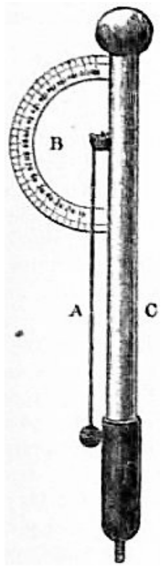

Fig. 1.—Henley's Electroscope.

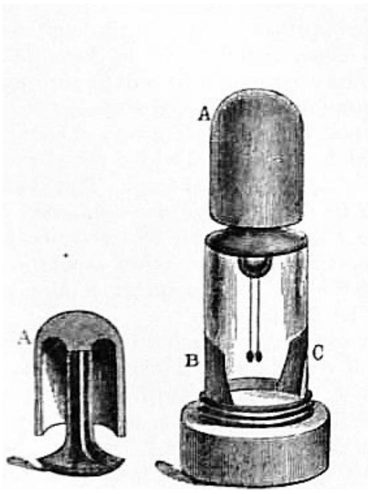

Fig. 2.—Cavallo's Electroscope.

Fig. 3.—Bennet's Electroscope.

The uses of an electroscope are, first, to ascertain if any body is in a state of electrification, and secondly, to indicate the sign of that charge. In connexion with the modern study of radioactivity, the electroscope has become an instrument of great usefulness, far outrivalling the spectroscope in sensibility. Radio-active bodies are chiefly recognized by the power they possess of rendering the air in their neighbourhood conductive; hence the electroscope detects the presence of a radioactive body by losing an electric charge given to it more quickly than it would otherwise do. A third great use of the electroscope is therefore to detect electric conductivity either in the air or in any other body.

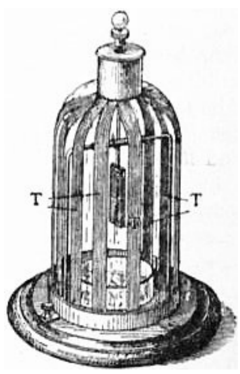

Fig. 4.—Gold-Leaf Electroscope.

To detect electrification it is best to charge the electroscope by induction. If an electrified body is held near the gold-leaf electroscope the leaves diverge with electricity of the same sign as that of the body being tested. If, without removing the electrified body, the plate or knob of the electroscope is touched, the leaves collapse. If the electroscope is insulated once more and the electrified body removed, the leaves again diverge with electricity of the opposite sign to that of the body being tested. The sign of charge is then determined by holding near the electroscope a glass rod rubbed with silk or a sealing-wax rod rubbed with flannel. If the approach of the glass rod causes the leaves in their final state to collapse, then the charge in the rod was positive, but if it causes them to expand still more the charge was negative, and vice versa for the sealing-wax rod. When employing a Volta condensing electroscope, the following is the method of procedure:—The top of the electroscope consists of a flat, smooth plate of lacquered brass on which another plate of brass rests, separated from it by three minute fragments of glass or shellac, or a film of shellac varnish. If the electrified body is touched against the upper plate whilst at the same time the lower plate is put to earth, the condenser formed of the two plates and the film of air or varnish becomes charged with positive electricity on the one plate and negative on the other. On insulating the lower plate and raising the upper plate by the glass handle, the capacity of the condenser formed by the plates is vastly decreased, but since the charge on the lower plate including the gold leaves attached to it remains the same, as the capacity of the system is reduced the potential is raised and therefore the gold leaves diverge widely. Volta made use of such an electroscope in his celebrated experiments (1790-1800) to prove that metals placed in contact with one another are brought to different potentials, in other words to prove the existence of so-called contact electricity. He was assisted to detect the small potential differences then in question by the use of a multiplying condenser or revolving doubler (see Electrical Machine). To employ the

electroscope as a means of detecting radioactivity, we have first to test the leakage quality of the electroscope itself. Formerly it was usual to insulate the rod of the electroscope by passing it through a hole in a cork or mass of sulphur fixed in the top of the glass vessel within which the gold leaves were suspended. A further improvement consisted in passing the metal wire to which the gold leaves were attached through a glass tube much wider than the rod, the latter being fixed concentrically in the glass tube by means of solid shellac melted and run in. This insulation, however, is not sufficiently good for an electroscope intended for the detection of radioactivity; for this purpose it must be such that the leaves will remain for hours or days in 240
a state of steady divergence when an electrical charge has been given to them.

In their researches on radioactivity M. and Mme P. Curie employed an electroscope made as follows:—A metal case (fig. 5), having two holes in its sides, has a vertical brass strip B attached to the inside of the lid by a block of sulphur SS or any other good insulator. Joined to the strip is a transverse wire terminating at one end in a knob C, and at the other end in a condenser plate P′. The strip B carries also a strip of gold-leaf L, and the metal case is connected to earth. If a charge is given to the electroscope, and if any radioactive material is placed on a condenser plate P attached to the outer case, then this substance bestows conductivity on the air between the plates P and P′, and the charge of the electroscope begins to leak away. The collapse of the gold-leaf is observed through an aperture in the case by a microscope, and the time taken by the gold-leaf to fall over a certain distance is proportional to the ionizing current, that is, to the intensity of the radioactivity of the substance.

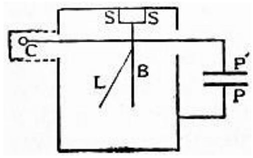

Fig. 5.—Curie's Electroscope.

A very similar form of electroscope was employed by J.P.L.J. Elster and H.F.K. Geitel (fig. 6), and also by C.T.R. Wilson (see *Proc. Roy. Soc.*, 1901, 68, p. 152). A metal box has a metal strip B suspended from a block or insulator by means of a bit of sulphur or amber S, and to it is fastened a strip of gold-leaf L. The electroscope is provided with a charging rod C. In a dry atmosphere sulphur or amber is an early perfect insulator, and hence if the air in the interior of the box is kept dry by calcium chloride, the electroscope will hold its charge for a long time. Any divergence or collapse of the gold-leaf can be viewed by a microscope through an aperture in the side of the case.

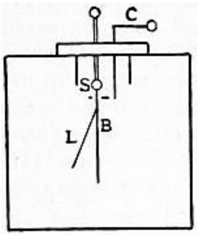

Fig. 6.—Elster and Geitel Electroscope.

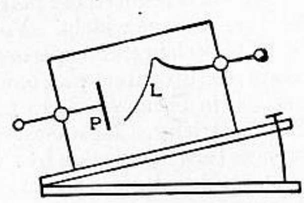

Fig. 7.—Wilson's Electroscope.

Another type of sensitive electroscope is one devised by C.T.R. Wilson (*Proc. Cam. Phil. Soc.*, 1903, 12, part 2). It consists of a metal box placed on a tilting stand (fig. 7). At one end is an insulated plate P kept at a potential of 200 volts or so above the earth by a battery. At the other end is an insulated metal wire having attached to it a thin strip of gold-leaf L. If the plate P is electrified it attracts the strip which stretches out towards it. Before use the strip is for one moment connected to the case, and the arrangement is then tilted until the strip extends at a certain angle. If then the strip of gold-leaf is raised or lowered in potential it moves to or from the plate P, and its movement can be observed by a microscope through a hole in the side of the box. There is a particular angle of tilt of the case which gives a maximum sensitiveness. Wilson found that with the plate electrified to 207 volts and with a tilt of the case of 30°, if the gold-leaf was raised one volt in potential above the case, it moved over 200 divisions of the micrometer scale in the eye-piece of the microscope, 54 divisions being equal to one millimetre. In using the instrument the insulated rod to which the gold-leaf is attached is connected to the conductor, the potential of which is being examined. In the use of all these electroscopic instruments it is essential to bear in mind (as first pointed out by Lord Kelvin) that what a gold-leaf electroscope really indicates is the difference of potential between the gold-leaf and the solid walls enclosing the air space in which they move.[1] If these enclosing walls are made of anything else than perfectly conducting material, then the indications of the instrument may be uncertain and meaningless. As already mentioned, Faraday remedied this defect by coating the inside of the glass vessel in which the gold-leaves were suspended to form an electroscope with tinfoil (see fig. 4). In spite of these admonitions all but a few instrument makers have continued to make the vicious type of instrument consisting of a pair of gold-leaves suspended within a glass shade or bottle, no means being provided for keeping the walls of the vessel continually at zero potential.

See J. Clerk Maxwell, *Treatise on Electricity and Magnetism*, vol. i. p. 300 (2nd ed., Oxford, 1881); H.M. Noad, *A Manual of Electricity*, vol. i. p. 25 (London, 1855); E. Rutherford, *Radioactivity*.

(J. A. F.)

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1 See the English translation by the Gilbert Club of Gilbert's *De magnete*, p. 49 (London, 1900).

1 See Lord Kelvin, "Report on Electrometers and Electrostatic Measurements," *Brit. Assoc. Report* for 1867, or Lord Kelvin's *Reprint of Papers on Electrostatics and Magnetism*, p. 260.