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Understanding
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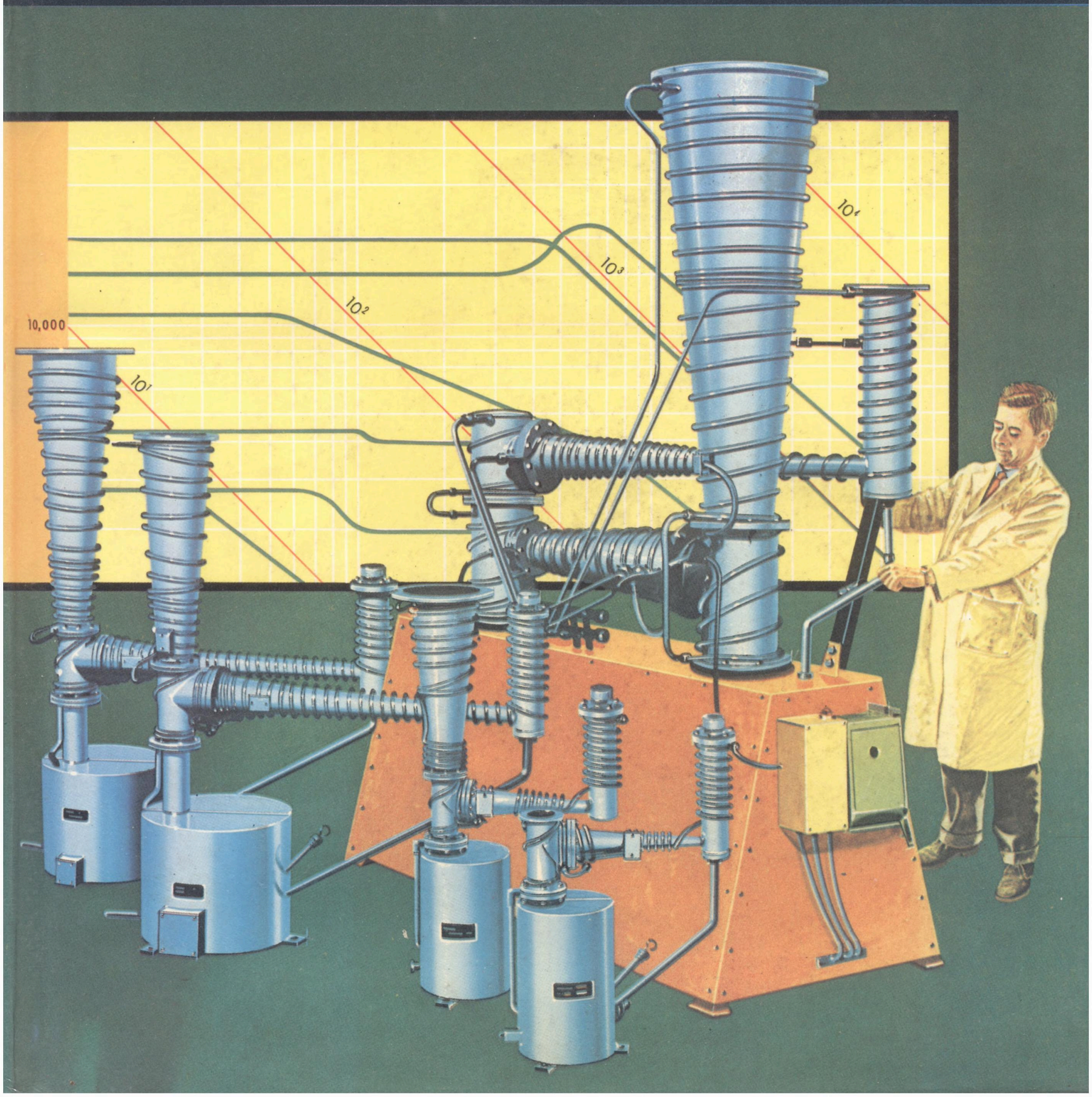

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UNDERSTANDING SCIENCE CLASSIFIED CONTENTS

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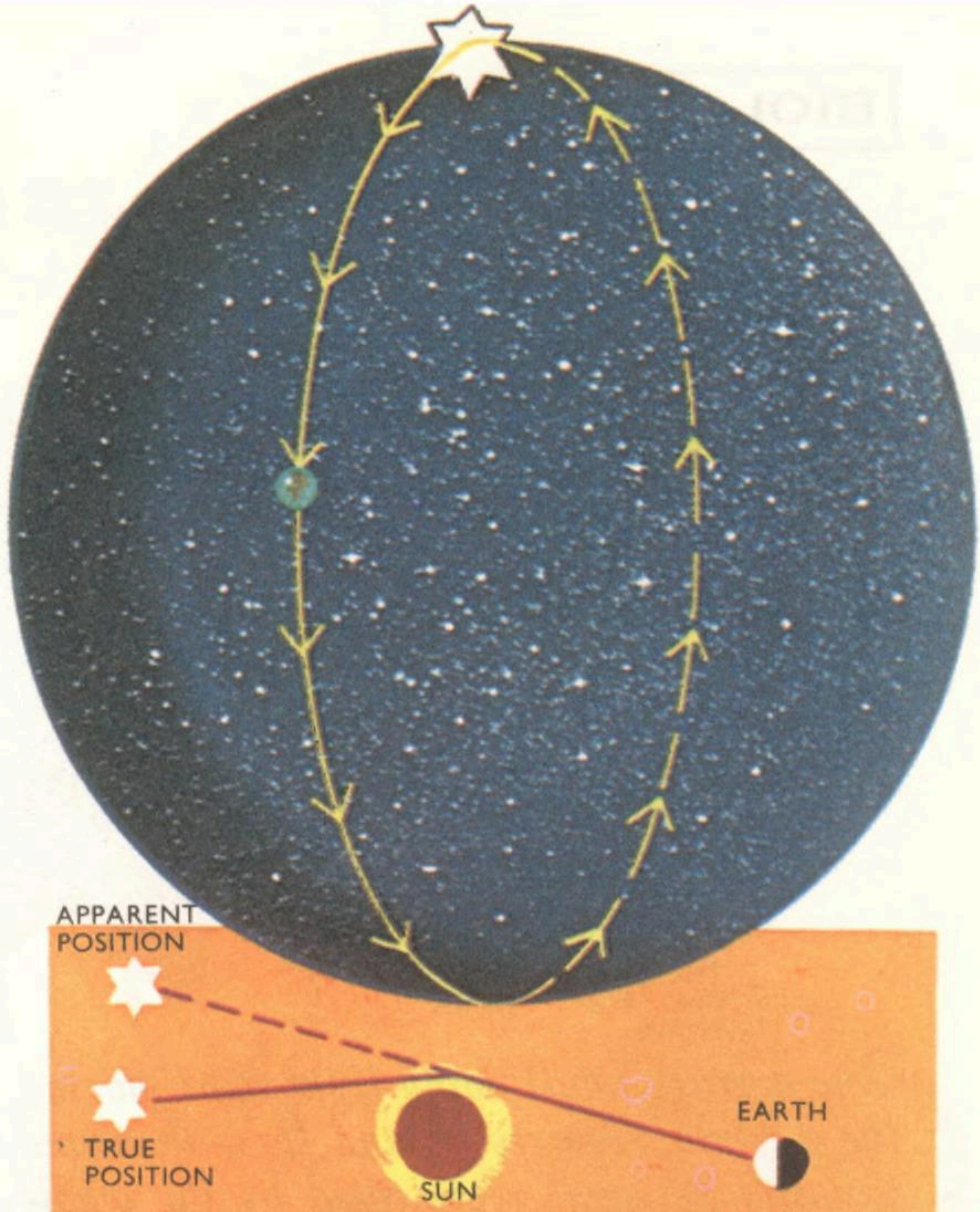
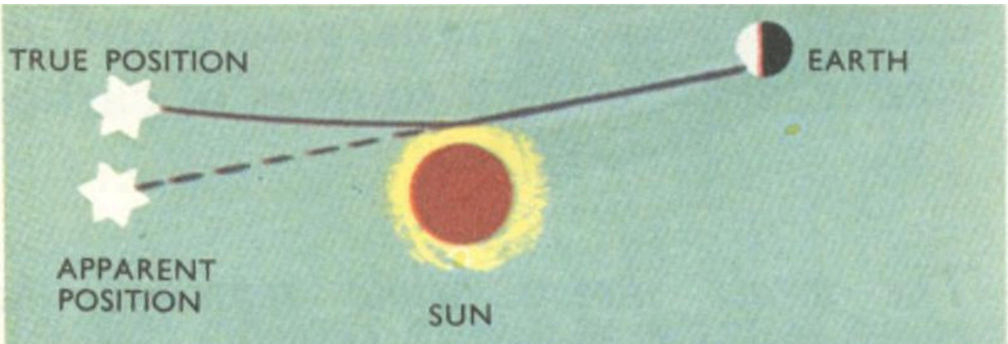
| ASTRONOMY T he Curvature of Space HAT is the shape of the Universe ? Is it cubical, spherical, or completely unbounded, stretching out to infinity ? All the information we have about the extremities of the Universe comes from the light (and radio waves) radiated by distant galaxies. Light seems to reach the earth from all directions, so perhaps the Universe is symmetrical, either spherical or infinite. But the Universe is neither of these. It cannot be represented fully by any three-dimensional geometrical figure. The outside boundary of the Universe cannot be pictured because light is not bringing its information in straight lines. All the space inside the boundary is curved. Space is not three-dimensional, like a building block or a sphere. It is fourdimensional, and the fourth dimension is time. It appears in equations governing the properties of space, but there is no way of picturing it. Space is curved and _ distorted because it contains matter — all the billions of billions of stars and galaxies in the Universe. Light is affected by the gravitational forces exerted by matter in space. Over long distances, light travels in curves instead of straight SUN D EARTH The uncurved universe. Time is not needed to define it—it could be pictured in three dimensions. The universe would not be curved if light travelled in straight lines. lines. Our own Sun is not a particularly massive star, but even it can produce a detectable bending in light coming from a distant star to the Earth, if the light passes within a few degrees of the Sun. The direction of bending observed seems to suggest that light bends inwards. A light ray starting out from any point in the Universe is, on the whole, attracted in towards the centre all the time. Eventually, after being attracted inwards by all the matter in the Universe, the light ray will eventually arrive back at its starting point. It is like travelling in a straight line away from any point on Earth. The ‘straight line’ turns out to be a curved path around the Earth’s surface. Every 24,000 miles (the Earth’s circumference) the path arrives back at its starting point, having formed a great circle. The curvature of space can be pictured by the odd behaviour of light — in particular by the velocity of light. Velocity is distance divided by time. An illustration including the behaviour of the velocity of light also includes the dimension of time (which is impossible to include in any purely spatial diagram). If light is unaffected by matter, and always travels in straight lines (i.e. at the same velocity) space is undistorted and uncurved. This can be pictured as a flat, two-dimensional surface (this is a

way of saving the third dimension until it is really needed). If light doubles back on itself on a - great circle route around the Universe, the two dimensional diagram turns into a three-dimensional sphere. The paths of light are circles around the sphere. The light changes its direction so its velocity changes (velocity takes into account direction as well as speed). Albert Einstein's relativity theories are all tied up with the behaviour of the velocity of light. In his General Theory of Relativity (1916) Einstein showed what would happen if light interacted with matter. There were three possibilities in his equations: either light was unaffected, in which case the Universe would be 'flat', or light could be bent either inwards or outwards. Both these second and third possibilities would produce a curved, four-dimensional space. But if light curved outwards instead of inwards, the diagram would become saddleshaped and the curves hyperbolae instead of circles . Light rays would fly outwards and never return to their starting point. Present experimental evidence seems to favour the inwardscurving space.

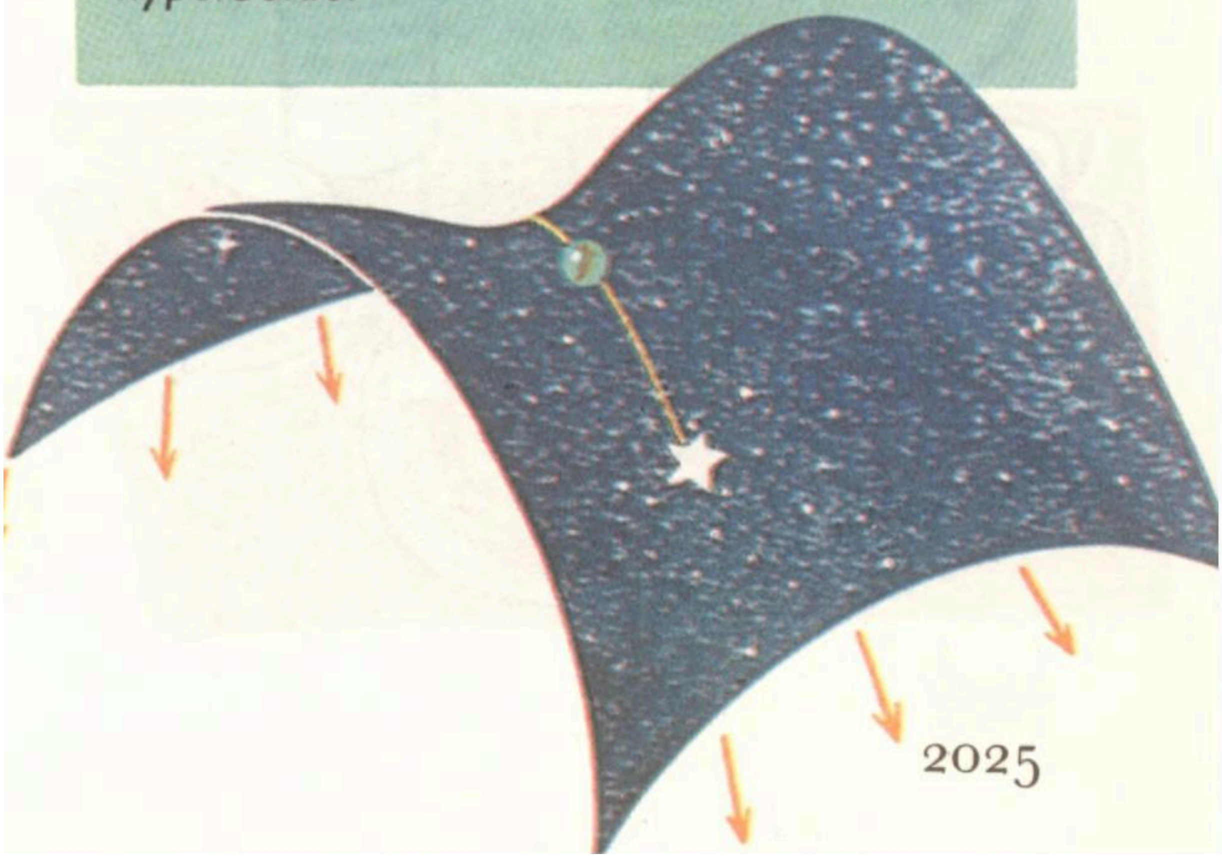


Space is thought to curve inwards, like a sphere. In other words, light, our only means of telling the nature of space, is bent in towards matter.

space, is bent in towards matter.



The universe could be represented like this if light curved outwards instead of inwards. The paths of light rays are *hyperbolae*.



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| BIOLOGY | The average adult doing light work during the daytime requires around three thousand Calories per day, represented by the whole circle. The diagram shows roughly how many Calories each major activity consumes. 1 Calorte =1 kilo calorie HE activities of the body are energy — consuming processes. The energy needed is obtained by the breakdown of the digested food materials by the action of enzymes within the cells of the tissues. Of prime importance as energy supplies are carbohydrates — starch, sugars and the like, of the other two basic food materials fats are more important as energy sources than proteins. The latter are primarily required for building new tissues and for repairing damaged or worn tissues, though under conditions of great stress — such as during serious illness — they may be respired in large quantities as can be seen by the wastage of muscle tissue. Besides fats, carbohydrates and proteins, a balanced diet must contain traces of vitamins, mineral salts and, of course, water. Rich sources of carboThe meals shown below represent the approximate Calorie intakes of a typical day's diet and the Calorie intake at mealBREAKFAST ABOUT 750 K. CALORIES The Quality of Food hydrate are cereals, bread, milk, beans, peas, raisins, milk chocolate, syrup, jam and cakes. Fat-rich foods include butter, margarine, cheese, cooking fats, bacon, pork, peanuts and plain chocolate. Rich sources of protein are eggs, fish, meat, beans and peanuts. The diagrams illustrate the varying proportions of the three basic food materials (fats, proteins, and carbohydrates) in various foods and also the number of calories that they yield per ounce. Foods such as sugar, which is all carbohydrate, are good energy suppliers, whereas milk and eggs are the best sources in animal protein. Foods rich in animal protein are better suppliers of proteins than plants because they contain a greater proportion of first class protein. Cereals are chiefly energy suppliers though wheat, for example, may contain as much as 14 per cent protein. Carbohydrates do and should form the greater part of the diet. When the diet is low in carbohydrate, fat is respired as the chief alternative source of food, but this eventually leads to acidosis — an extremely high acidity of the body fluids. Fats are very economical storage materials, however. For every gram of fat that is completely respired, 9°3 kilo calories are yielded. The figure for 1 gram of carbohydrate is only 4-1 k. calories and for protein 4:1 k. calories. times and in between meals during the day. The place mats are drawn to scale and

MID-MORNING BREAK ABOUT 100 K.. CALORIES Carbohydrates, which are made up of carbon, oxygen and hydrogen atoms, can be divided into two classes — the sugars and the polysaccharides. Sugars are crystalline solids, sweet to the taste and soluble in water. The simplest of them — called monosaccha-. rides — either have the chemical formula C,H,,O; (the pentose sugars such as xylose) or C,H,,0, (the hexose sugars such as glucose). Slightly more complicated are the disaccharides, which have their molecules made up of two molecules of monosaccharide (e.g. sucrose). Also found are trisaccharides and tetrasaccharides made up of three and four monosaccharide molecules respectively. Polysaccharides are usually insoluble powders (e.g. starch, agar). Really they are gigantic sugar molecules for they are made up of hundreds, even thousands of monosaccharide molecules. The digestion of carbohydrates is simply the breaking down of polysaccharides and disaccharides into basic monosaccharide molecules which can be absorbed into the blood stream. Fats are esters (organic salts) produced by fatty acids combining with alcohols — usually glycerol. Like carbohydrates, they are also made of carbon, oxygen and hydrogen atoms. During digestion, fats are broken down into their components i.e. fatty acids and glycerol. As well as carbon, oxygen and hydrogen, proteins contain nitrogen, sulphur and sometimes phosphorus. During digestion proteins first break down into fragments called peptides. These peptides, in turn break down into the basic protein units, the amino acids. indicate that more Calories are obtained from the main meal in the day than any ABOUT 950 K. CALORIES

Carbohydrates, which are made up of carbon, oxygen and hydrogen atoms, can be divided into two classes – the *sugars* and the *polysaccharides*.

Sugars are crystalline solids, sweet to the taste and soluble in water. The simplest of them – called *monosaccharides* – either have the chemical formula $C_6H_{10}O_5$ (the pentose sugars such as xylose) or $C_6H_{12}O_6$ (the hexose sugars such as glucose). Slightly more complicated are the *disaccharides*, which have their molecules made up of two molecules of monosaccharide (e.g. sucrose). Also found are *trisaccharides* and *tetrasaccharides* made up of three and four monosaccharide molecules respectively.

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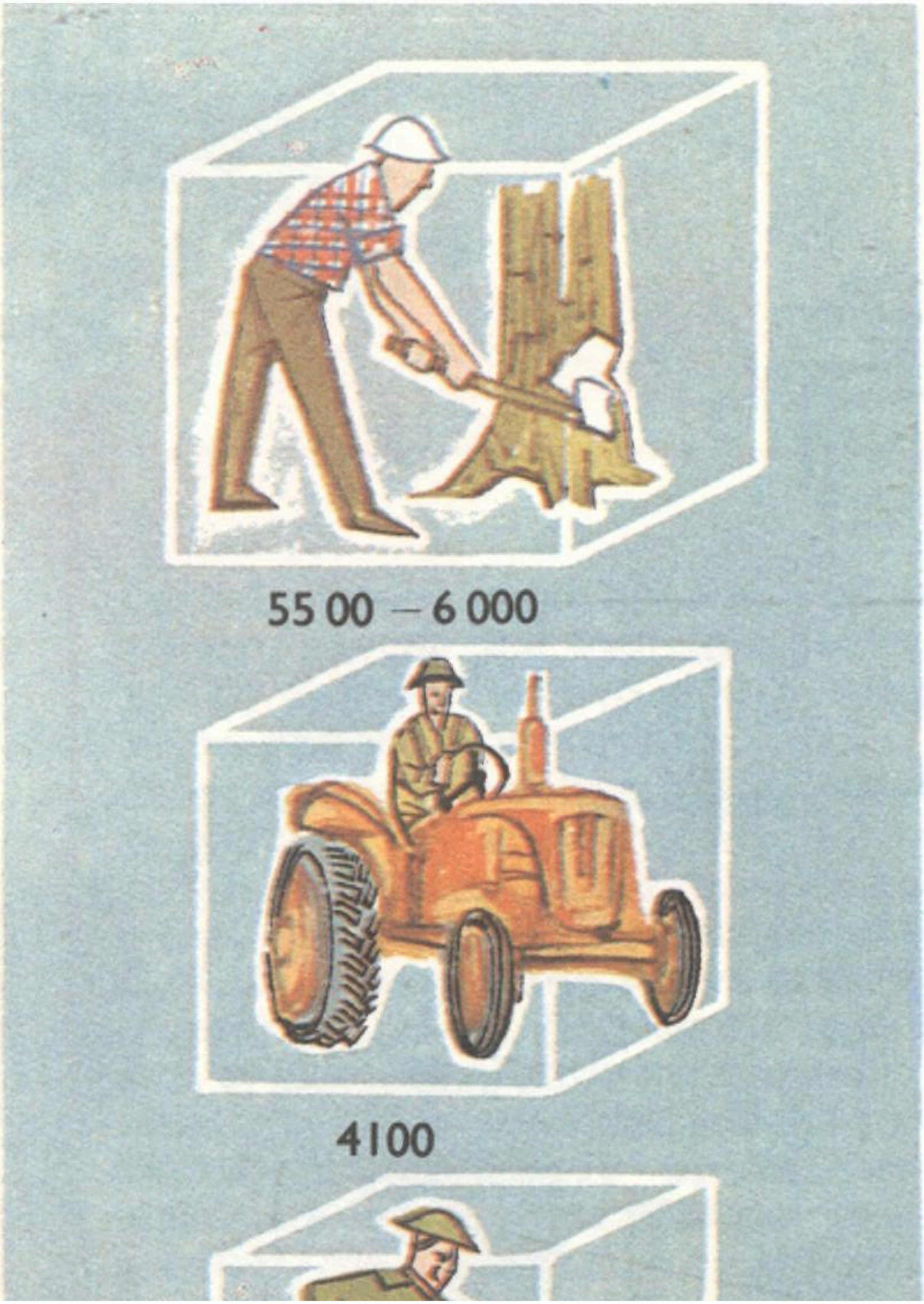
As well as carbon, oxygen and hydrogen, proteins contain nitrogen, sulphur and sometimes phosphorus. During digestion proteins first break down into fragments called *peptides*. These peptides, in turn break down into the basic protein units, the *amino acids*.

The calorific requirements of individuals vary considerably depending on both physiology and occupation. Thus a person doing sedentary work will require a lower intake than a manual worker. An athletic person may require less than an obese person in a similar occupation because his body operates more efficiently than the latter's. Waste substances produced by the tissues are removed more rapidly by a more efficient circulatory system. The calorific requirements of a woman are generally less than those of a man, primarily because her surface area is much less than that of a man. Special physiological conditions also affect dietary requirements. Thus a mature woman requires a diet rich other meal. Note that meals where considerable quantities of starchy and sweet AFTERNOON 'TEA' ABOUT 60 K. CALORIES 88 GM PROTEIN/DAY 56 GM PROTEIN/DAY 94 GM PROTEIN, DAY World map showing the average daily Calorie intake per head of the population in seven main regions — this is represented by the proportionate circles — and also the daily protein intake in grams — shown by proportionate cubes. The circles also show proportion of calories derived from carbohydrate. in iron to compensate for the loss of this element in blood lost during menstruation. Pregnancy makes great demands on her resources of calcium, phosphorus and protein, and the production of milk whilst suckling a child makes excessive demands on fat metabolism. The columns show the relative amounts of energy released by the complete combustion of 1 gm. of fat, carbohydrate and protein. Fat supplies more than twice as much energy per unit of weight as carbohydrate and protein. foods (e.g. bread, potatoes, sugar) are consumed have high Calorie values. The EVENING MEAL ABOUT 740 K. CALORIES relative 'amounts of energy supplied by one ounce of various foods is shown in the diagrams overleaf. SUPPER ABOUT 400 K. CALORIES 2027

CALORIFIC VALUES OF COMMON FOODS THE CALORIES ARE KI LORIES BUTTER Fen eg Sy 211 CALS =) PER OUNCE WHITE BREAD 73 CALS PER OUNCE CHEESE 117 CALORIES PER OUNCE POTATOES 21 CALORIES PER OUNCE Cie % CARBOHYDRATE 2028 BISCUITS 126 CALS PER OUNCE MILK 17 CALS PER OUNCE 71 CALORIES PER OUNCE . EGGS 45 CALS PER OUNCE ¥ ry, Ysa 99 CALORIES PER OUNCE LIVER 40 CALORIES PER OUNCE CABBAGE 7 CALORIES PER OUNCE % PROTEIN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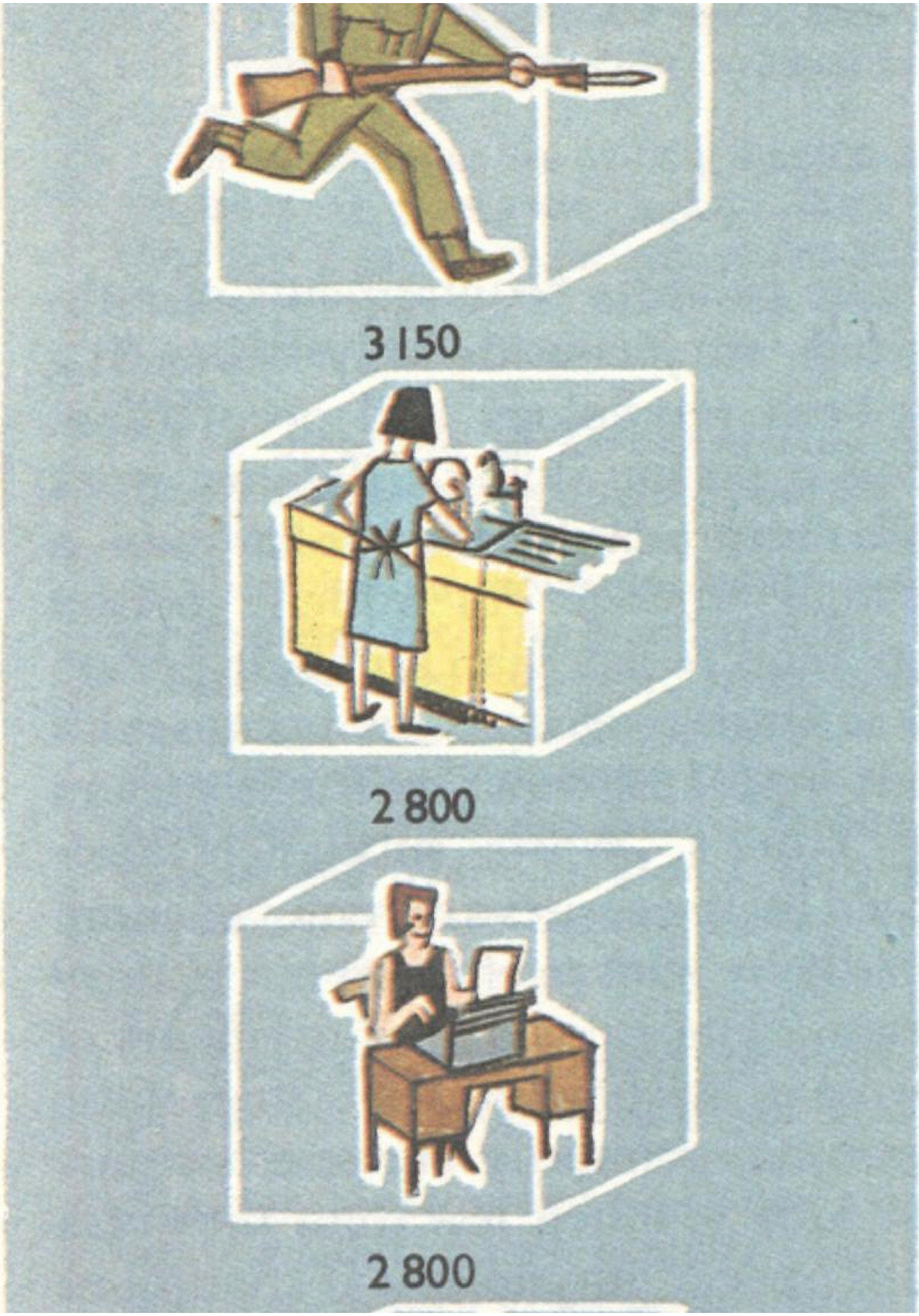
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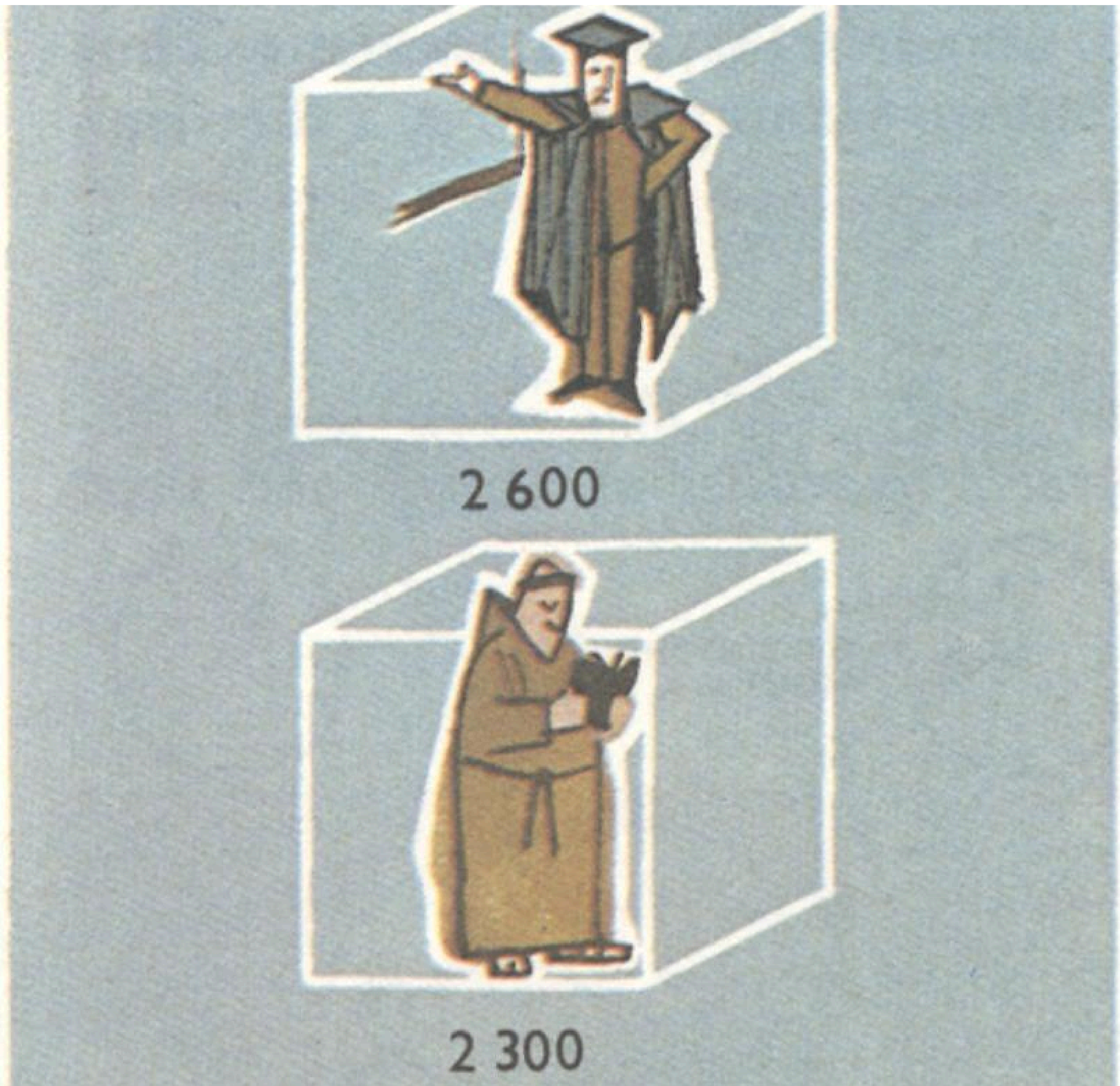
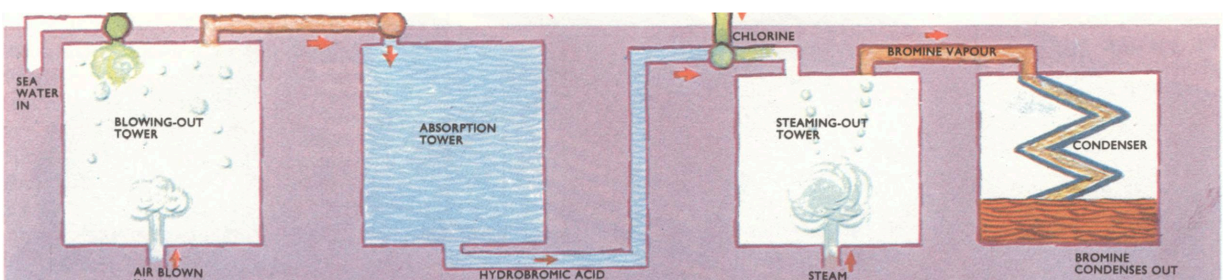


Approximate number of Calories per day required by people in different occupations.

CHLORINE AS INORGANIC CHEMISTRY The Heavy Halogens IKE the other halogens the two heaviest common members of the family, bromine and chlorine, are not found free in nature. They are much too reactive, the elements being strongly electro-negative. This means that bromine and chlorine atoms are very eager to accept a single electron from the atoms of another element, and so are able to take part readily in chemical reactions. The heaviest halogen of all, astatine, is not found in nature at all. It is a comparatively shortlived radio-active element, formed when bismuth is bombarded with alpha particles. Bromine Bromine is a reddish-brown liquid that boils at 58 -8°C to give off a dark brown vapour. It has a pungent odour and attacks the eyes and the throat, as well as the skin. It is found in seawater and in natural salt deposits in combination with sodium, calcium, potassium and magnesium. An important industrial method of preparation is to displace the bromine from the metallic salt using chlorine gas. Salt is extracted from sea water, and the remaining liquor is passed down a reaction tower, whilst chlorine gas flows up it in the opposite direction. The chlorine reacts with the magnesium bromide in the liquor: MgBr, + Cl = MgCl, + Bry This reaction provides an example of the greater chemical activity of the lighter members of the halogen family. Although the magnesium and bromine atoms are fairly tightly bound in magnesium bromide, the greater electronegativity of the chlorine atoms is sufficient to make them oust the bromine atoms from the compound. Bromine is used in the manufacture of photographic film — silver bromide is a light-sensitive material, and the bromides of sodium and potassium are used as sleeping drugs. But the main use of bromine is in the form of ethylene dibromide (C.H,Brg), as an antiknock additive in petrol. lodine ‘Iodine? as commonly known, is a brown liquid, used for dressing wounds. In fact, this is not pure iodine, but is a tincture formed by dissolving the element in alcohol. The pure element itself is a black lustrous metallic-looking solid that melts at 113°C, At 187°C it boils to give off a deep violet-coloured vapour. It is found in small quantities in sea water and in sea plants, particularly sea weed, but the main source is the nitrate beds in Chile, where it is found in the form of sodium iodate, NaIOs. Sodium iodide is used in photographic film, and iodine salts are essential in the diet, for growth and metabolism in human beings. Although the iodine atom ranks as being highly reactive compared with many atoms of other elements, it is less reactive than the other common halogens. It is less electronegative than any of the other halogens and is a weaker oxidizing agent. It is possible to get a good idea of the reactivity of a chemical substance by measuring the amount of heat given out when it reacts. When iodine combines with the metals it gives out less heat than the other halogens in similar reactions. In fact the metal sometimes has to be heated before it will combine, demonstrating its comparatively low reactivity. The Heaviest Halogen - Astatine All the halogens have seven electrons in the outer shell of their atoms, and the heavier the atom, the less reactive it is. Their physical properties are graded with increase in mass. Fluorine, the lightest halogen, is a very reactive gas. Iodine is a near-metallic solid. Its atomic weight is 126 -g and it was thought that it was the heaviest existing halogen. Then, in 1940, the element astatine was discovered, and found to be a member of the halogen family. Its atomic weight is 211, and it is a short-lived radioactive element, whose properties are not yet fully known. However, it appears to resemble iodine in many ways, having an even more metallic appearance. Like iodine it can be taken up by the thyroid gland where the dangerous radiation it emits can destroy living tissue. At normal temperatures pure iodine is a black metallic-looking solid. It can be detected in solution by the very sensitive starch test. Starch paper turns a deep blue in the presence of todine. STARCH PAPER CRYSTAL FORMS OF IODINE 2029



Extraction of Bromine from Brine Liquid. Chlorine is passed into the brine and replaces chlorine from the bromide salts. The bromine is blown out with an air blast and the bromine-laden air is mixed with sulphur dioxide. Hydrogen bromide is formed and this is mixed with chlorine which again releases bromine. This is steamed-out in the steaming out tower and passes on to the condenser.

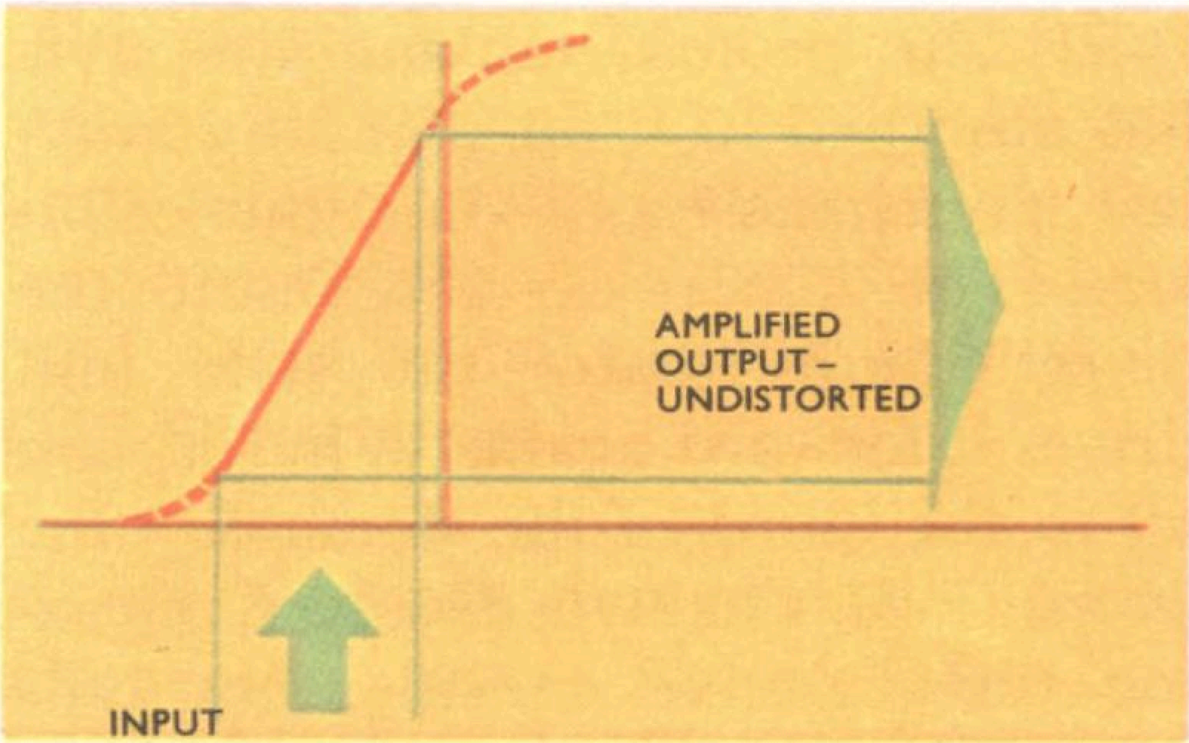
| TECHNOLOGY | AUTOPILOT CONNECTED TO GLIDE SLOPE AND OGALISER"SIGNALS i St te Stee % Oss a ar,, >. > % ‘ ep I & RTI a nr ae Me ae iy tan w Sina, SK ea ae tian a fing te of an nage mak = Eling | ae ae Be Riise n tells the autopilot where ing et. Fhe * as on marksthe cect OR, is phe mnviey’ hreshol how the distance, to fs ns é > » oy . Pr = ae BLIND LANDING SYSTEMS HE idea of flying in an aeroplane which is controlled entirely automatically may seem rather frightening. But airliners already are flown tremendous distances on the autopilot (called ‘George’). ‘George’ flies the plane for him, while the pilot has only to monitor, or keep an eye on, the plane’s progress. If George goes wrong the pilot can quickly switch him off and take full control. At present, all landings of aircraft (apart from occasional experimental flights on which, of course, no passengers are carried) are made under the control of the pilot, often assisted by a controller sitting in the airport control tower watching the plane on special radar screens. He can see if anything is seriously wrong with a plane’s approach and warn the pilot by radio. A modern transatlantic jet, weighing perhaps 100 tons, lands at about 160 m.p.h. Considerable precision is necessary to ensure it arrives at the beginning of the runway at the right height and speed. As the plane reaches the runway, the rate of descent is gradually reduced, until, as it touches the ground it is hardly descending at all. This manoeuvre is called ‘flare-out’. All the while the plane must of course be kept flying in a straight line with the runway, 2030 despite any crosswind which may be blowing. To compensate for crosswind, the plane has to be flown with its nose pointing slightly up-wind. If it landed like this, as soon as the wheels gripped, there would be a tremendous sideways strain on the undercarriage and the plane would tend to veer off the runway in the direction in which it was pointing (i.e. up-wind), so immediately before the wheels touch, the pilot applies rudder to straighten up. This is called kicking-off drift. Such a landing calls for great concentration on the part of the pilot, and there is always the possibility of error. In fact, of all flying incidents, nearly half occur during the final stages of the approach (even so, the accident rate is very low, about 1 in 150,000). If visibility is bad, the pilot’s problem becomes extremely difficult, due to the short time he has in which to assess the situation and when visibility is less than 400 yards aircraft cannot land and must be diverted. Full automatic landing systems would not require the pilot to see

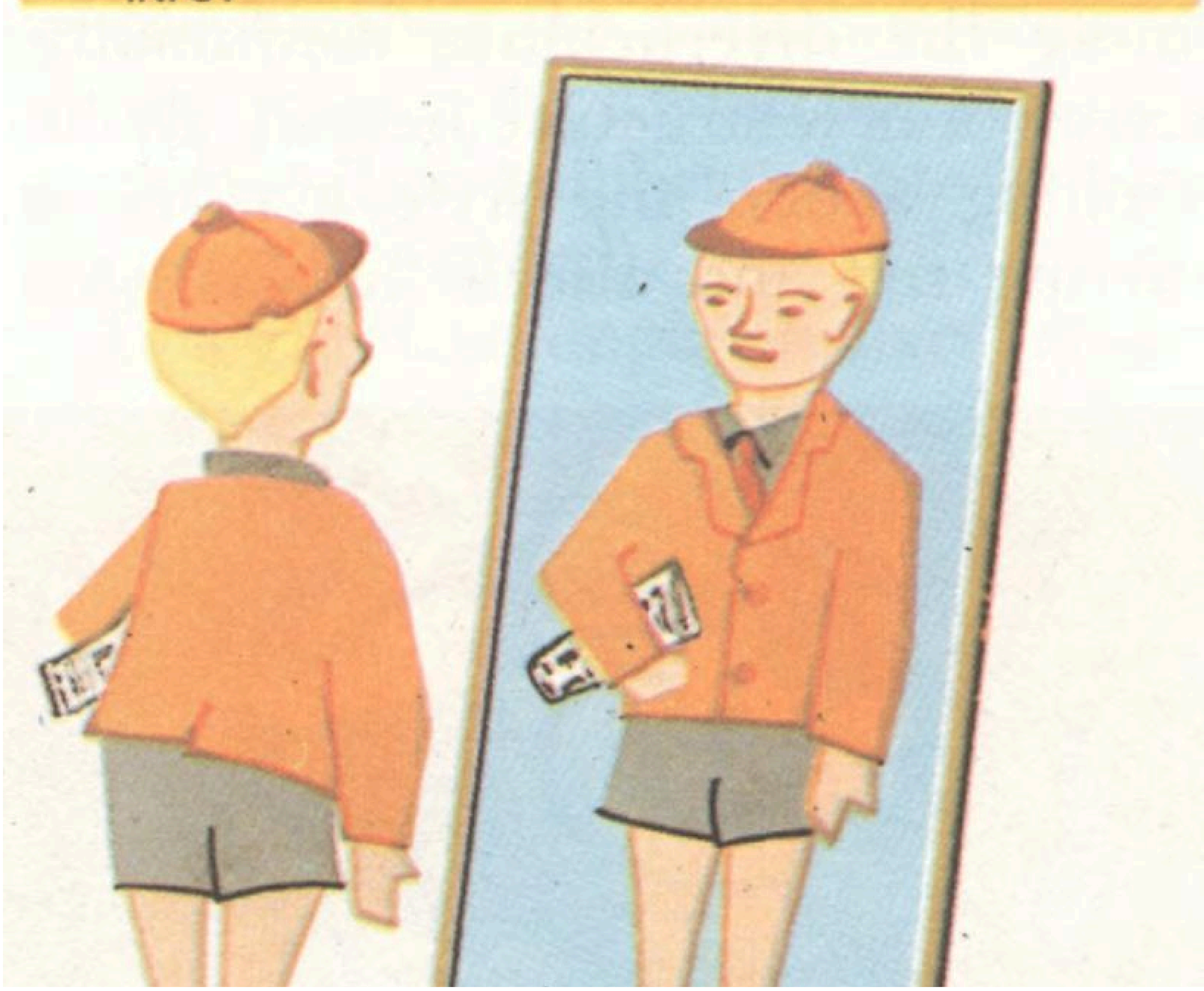
any more than the instruments on_ his control panel (which is all he needs to see when flying high up on course in the air corridors, with or without ‘George’). Two stages of blind landing systems are at present being developed. Autoflare in which the plane is automatically flown to just a few feet above the runway threshold, leaving the pilot to kick-off drift and land, and Autoland, in which the entire approach, flareout and landing is done automatically. Autoflare still requires sufficient visibility for the pilot to see enough of the runway to steer the plane down it. Autoland could be undertaken blind. The control of an aircraft’s landing can be split into two. First, the control necessary to keep it in alignment with the runway and, second, the control to keep it on its proper glide path so as to arrive at the beginning of the runway at the best landing speed. One method of Autoflare makes use of the already installed Instrument Landing System (ILS) groundequipment. Here, very accurate radio beams transmitted from equipment on the airfield define the runway centre line and the glide path. One transmitter provides a signal down the extended runway centre line, this is called the localizer. Another defines a 3 degree glide path which intersects the runway at the touch-down point. These signals may be picked up 7 to 10 miles from touchdown and in the ordinary ILS an instrument in the aircraft will indicate to the pilot the aircraft’s position relative to the centre line and whether it is above or below the glide path. Marker transmitters located at two or three points on the approach path give a vertical radio beam which indicates the distance from touchdown. In blind landing systems the localizer, glide path and marker transmissions are received and decoded by the aircraft equipment and used to compute instructions which are fed to an autopilot (an elaborate' version of 'George'). The remaining requirement is an accurate indication of height when very near to the ground. This is provided by a radio altimeter mounted in the aircraft. Continuous radio signals are beamed at the ground and their reflections picked up by a receiver aerial. By measuring the time taken for the signal to reach the ground and return, the height of the aircraft can be computed. These altimeters are accurate to within 6 inches at an altitude of 10 feet and for automatic landings the zero must be adjusted for the length of the undercarriage legs. To commence an automatic landing the aircraft is guided by radar in the normal way to a point some 7 miles from the end of the runway, where it will intercept the ILS beams. At this stage the undercarriage and flaps are lowered to prevent unnecessary changes of trim later in the descent. The ILS signals are decoded, instructions computed and passed on to the autopilot and the descent begins from about 1500 feet. As it gets close to the ground the accuracy of the ILS glide path information decreases and at about 150 feet the radio altimeter switches out the ILS signals. From here to about 50 feet, pitch control (nose-up or nose-down) is by memory guidance. This information, which has been stored in an 'attitude-hold' computer, is the mean of all the pitch information collected over the earlier part of the glide path. At 50 feet the radio altimeter takes over for the automatic landing stage. Height information is fed to the autopilot and the rate of descent is made proportional to the aircraft's height. Thus, the lower the aircraft descends, the slower becomes the rate of descent. Consequently, a few inches above the runway it is hardly descending at all and so a smooth landing may be obtained. Autoflare takes the automatic sequence no further than this. The air craft should now be positioned over the runway threshold and all the pilot has to do is kick-off drift with the rudders, and land. Autoland requires a further refinement to ensure that the aircraft touches down on the runway unaided by the pilot. One possible method is the use of leader cables. Two electric cables buried either side of the runway may be energized to create a magnetic field. A device in the aircraft senses the resultant field strength and computes the aircraft's position between the two. This information is used to guide the aircraft down the centre line. Radio altimeter signals are used to govern the moment at The controls of a Hawker-Siddley Trident, a jet atr-liner designed to be used in the future with automatic landing equipment. The thumb-button on the pilot's control column is his "instinctive cut-out' which immediately gives him full control of the aircraft. INSTINCTIVE CUT- eS OUT BUTTON f which rudder must be applied to counter drift and so align the aircraft with the runway heading (pre-set on the gyro compass). The pilot merely monitors the equipment. The equipment used for blind landing systems is complex and liable (as all things are) to failure. Provision must be made to override faulty equipment in the event of its failing during a landing. The most promising system of reducing the risk of failure, is mulézplexing. In the Duplex system, the entire set of equipment is duplicated. If either set fails or issues a wrong signal oe HEIGHT e Flight Control Panel by which the pilot can instruct the Autopilot to control the aircraft through a blind landing. He can also use it at other times during flight, for example to fly to ('acquire') and stay at a particular height or speed. The three levers connect the autopilot to each axis of control. When the levers are fully back the autopilot cannot control the plane in that axis. The pilot may leave the autopilot to control, say, pitch, at a steady height and control the plane on the azimuth axis (direction) himself, for instance when he may have to make frequent changes of direction according to instructions he receives from Air Traffic Control. this is automatically detected because the two systems disagree. Both are then automatically switched out, leaving the plane in the pilot's control. In the Triplex system, suitable for total blind landing, where the pilot could not take control in the event of failure, there are three complete sets of equipment. This means that if one set fails, it can be identified as the odd one out (whereas in duplex it is not possible to decide which is at fault). So the Aulty circuit is switched out leaving the plane in the control of the other two. The chance of equipment's failing - can be accurately determined, and it is calculated that a full triplex system would have a failure rate of only 1 in 10 million — many times safer than a human pilot unaided. 2031

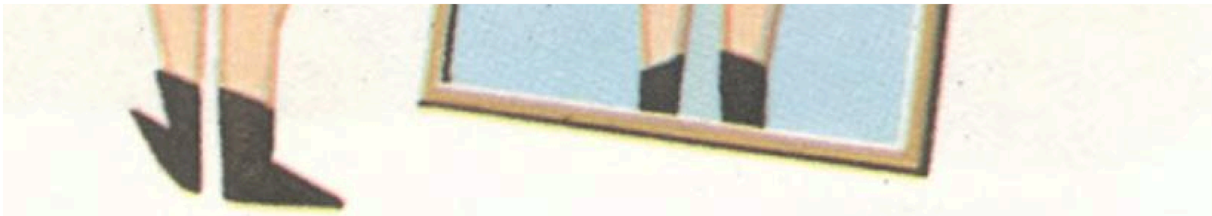
WATER PASSES POSTERIOR ADDUCTOR MUSCLE | BIOLOGY | RIGHT VALVE ANTERIOR ADDUCTOR MUSCLE ay The left valve and mantle lobe has been removed leaving the gill — plates and the palps exposed. A section of the gill plate has been ‘4 cut away, showing the structure of the gulls more closely. Insert: *\ MUSCULAR FOOT FILTER FEEDERS At a first inspection, the differences between a cockle and a snail seem to outnumber the likenesses. In fact, there hardly seems any resemblance between the two animals at all. Yet the snail (a gastropod) and the cockle (a bivalve or lamellibranch) are structurally very similar. They both belong to the Mollusca — one of the great divisions of the animal kingdom. Both have a hard protective shell, a mantle (a tent-like flap of tissue surrounding the body), a cavity within the mantle containing gills, and a powerful muscular foot. The ancient molluscs, ancestors of bivalves, gastropods and the other molluscan animals (cephalopods, scaphopods and amphineura) most probably crawled about in shallow water feeding off minute organisms which they lapped up using a horny, rasping In the simplest molluscs — ne peocbranichs — the gills remain simple. The ‘stalk’ gue $ these join to form two plates. In more advanced bivalves, each filament grows downwards and then upwards so that a descending and ascending off filaments on each side an section is formed. Two plates on each side are formed. W: inside through pores but! food particles are reta pe STRUCTURE OF GILLS IN VERY SIMPLE MOLLUSCS 2032 tongue or radula, Water percolating through the mantle cavity, washed against the gills providing oxygen. The gastropods have remained closer to this original way of life. They still move about looking for food, and possess a distinct head armed with a horny radula. But the bivalves have moved a long way from the ancestral mode of living. They have become filter feeders, Exclusively aquatic creatures, they waft tiny food particles into their mouths using tiny hairs or cilia usually arranged on the gills. The gills, increasing in importance, have also increased in size. The mantle cavity which houses them has become elongated, running down either side of the animal, often for almost the entire length. Bivalves are therefore usually long and compressed from side to side, ADVANCED STRUCTURE OF GILLS BY TISSUES detailed section of the gills showing cilia, pores and food grooves. BIVALVE MOLLUSCS The mantle itself has become divided into two lobes. Each lobe secretes a segment of shell, so the final structure is made of two valves — not, a single structure. But the mantle does remain slightly connected at the top (dorsal) surface. Here a ligament — an elastic strip made of organic material — is secreted, either on the inside or the outside of the hinge. Two muscles run between the valves, a forward anterior adductor and a backward posterior adductor. When the muscles contract, the valves close together but when the muscles relax the elastic ligament at the hinge causes the shell to gape. Bivalves, as filter feeders, do not really have a distinct head like snails. The mouth is at the front (anterior) end of the animal surrounded by two long bilobed palps (labial palps) but no sense organs are concentrated here and there is no rasping radula. Gill Structures In snails the gills are simple, consisting of a stalk giving off plume-like filaments — rather like feathers. But in the bivalves the filaments become joined up forming a solid or plate-like structure (lammellibranchiata in fact means, plate-gilled). In a very few simple bivalves (Protobranches) the gills remain confined to a cavity at the rear of the body and the filaments are short. In more advanced forms each filament becomes elongated and turns upwards. The gills are also extended so that they run practically the length of the body. Water is drawn into the shell at the rear (posterior) end where the mantle is ILAMENTS MAY BE INTERLOCKED often extended backwards to form two siphons. The lower siphon takes the water in. The water not only provides oxygen. Particles of food in the water are swept forward and downwards moving along the bottom of the gill filaments by rapidly beating of tracts of cilia. During the passage forward the particles become entangled in mucus. From the gills food particles are swept onto the labial palps about the mouth. The labial palps are also ciliated. Material may be swept into the mouth or alternatively rejected, passed back by cilia action along the base of the gills, and washed out through the upper of the two siphons. Sorting at the labial palps selects the finest particles, not necessarily the most nutritious. Large, coarse particles are rejected. Digestion Bivalves have no masticating radula and no salivary glands. Food particles are swept down a short gullet to the stomach. Cilia in the stomach further sort the particles and useless material is swept straight into the intestine. Some preliminary digestion does take place in the stomach. Digestive enzymes, for breaking down carbohydrates, are secreted there by a very strange method. Inside a diverticulum of the stomach is embedded a tough rod-like structure called a crystalline style. The protein making the style has absorbed in it for digesting amylase, glycogenase and sometimes cellulase. The action of cilia inside the diverticulum rotates the style and drives it forward against a hard plate set in the lining of the stomach — the gastric shield. The style dissolves, releasing the digestive enzyme, but cells lining the diverticulum continually secrete a new structure behind. By its rotating action the style acts as a stirring rod and also as a windlass; threads of mucus bearing food are drawn from mouth into stomach. Minute food particles are next taken via a pair of ducts to the digestive gland. The cells making this structure actually absorb the particles and digestion is completed inside them (intracellular digestion) . Three Groups of Bivalves Bivalve molluscs, like snails retain a muscular foot. The foot lies between the two sets of gills on the underneath (ventral) side of the body. Its shape and position is altered by muscular contraction and the structure can be protruded through the gape of the shell. In one group of bivalves — the socalled Normal bivalves — the foot is well developed and the creatures use it to move about at the surface (e.g. cockles) or in some instances to burrow (e.g. razor shells). The giant clams of the Pacific and Indian Oceans belong to this group. They may grow to be more than a yard in length. Food is not only obtained by filtering. Masses of small green unicellular algae grow in the tissues of the clam and sunlight is focussed upon them by numerous lenses. Some of the algae are periodically digested. The Sessile group of bivalves have, in general, lost their mobility in the adult animal. Instead they have become firmly attached to the surface. Modifications have accompanied this stationary mode of life. The ventral foot has tended to move more and more towards the front of the animal. Asa direct result the anterior adductor muscle has become small or is lost ; usually each valve becomes highly asymmetrical in shape. A byssus nearly always appears at some stage in the life history. A byssus is a mass of sticky, diverging threads arising from a pit at the back of the foot. In the mussel the byssus proDIAGRAMMATIC SECTION THROUGH BIVALVE SHOWING DIGESTIVE TRACT DETAIL OF BIVALVE STOMACH SHOWING GASTRIC SHIELD, STYLE AND DIGESTIVE GLAND A COMMON SEA SHORE SHELL PALLIAL LINE LEFT BY EDGE OF MANTLE MYTILUS THE MUSSEL HOLAS THE PIDDOCK. Top, Shell of Venus.Where the muscles abut against the shell, scars are left. The pallial line marks the edge of the mantle lobe. An insertion of the pallial line marks the position of the syphons. Middle, Mussels, sessile bivalves, have a reduced forward muscle, the scallops have lost it altogether. Below, Specialized burrowing forms. trudes through the gape in the valves. The pearl oyster has come to lie on its right valve and the byssus emerges through a notch. The common oysters have lost the foot and byssus altogether; the right valve becomes cemented to the ground. Pecten, the scallop, is structurally a sessile bivalve. The front adductor muscle has been lost by the forward migration of the foot. But the remaining muscle has become enlarged. By rapid contraction the valves are clapped together and water is violently discharged. This is an_ efficient mechanism for swimming and Pecten has given up its sessile habit. The last division of the bivalves is the deep-burrowers. Unlike the burrowers of the Normal group, the deep burrowers have become extremely modified to their mode of life. The valves are only weakly hinged; often they have completely lost their ligament and the valves permanently gape. The siphons are enormous, often dwarfing the rest of the body. 2033

ELECTRONICS AMPLIFICATION— CLASSES A, B& C

"THE grid of an electronic valve is a fine wire spiral encircling the cathode, and it is the controlling influence over the current flowing through the valve. It is nearer the cathode (the electron emitter) than the anode. A signal voltage of 2 or 3 volts applied to the grid has as much effect as a signal voltage of around 50 volts applied to the anode. The grid voltage varies by a small amount, and the anode voltage varies by a large amount. The grid is usually the input and the anode is part of the output circuit of the valve. The output signal is bigger than the input signal. The valve amplifies. But the valve's output is not necessarily a perfect, enlarged replica of the input. Valves sometimes distort the signal. The valve characteristics are two sets of curved lines (graphs) which show the variation of anode current with grid voltage, and anode current with anode voltage. Although the middle part of a triode valve's characteristic is practically straight, it curves at the ends. Over the straight portion, the output is proportional to the input, and the valve does not distort the signal. If the anode. voltage and the grid voltage are set so that the valve operates over the curved portion of its 2034 characteristic, however, the signal becomes distorted. The three ways of operating a valve over different parts of its characteristics are called Classes A, B and C. Class A A Class A valve amplifier is operated entirely in the straight portion. It produces a distortion-free output and is used as an audio-frequency amplii} CURRENT SWING | GRID VOLTAGE VOLTAGE Part of the characteristic 1s a straight line. The maximum variation in the grid voltage, for completely undistorted output, is shown by the horizontal arrow. If the voltage swing is greater than this, it encroaches into the curved parts of the characteristic, and distortion may set in. The grid voltage 1s nearly always made negative (with respect to the cathode). If it is positive, it attracts some of the negatively-charged electrons, which would otherwise have gone to the anode. This produces an unwanted grid current. The maximum current swing in the output is shown as the 'reflection' from the characteristic. fier. The one disadvantage is that the swing, to-and-fro, of the output voltage is restricted. The reason for this is that the valve is initially set with its grid voltage (grid bias) right in the middle of its specified values. This makes the valve operate in the middle of the straight portion of the characteristic curve. During half of the signal cycle the grid voltage is greater than the bias voltage, and during the other half it is less. But the swing is still restricted to the straight part of the characteristic. The input swing is limited, so the output swing is also limited. Class B A Class B amplifier produces a large, undistorted swing in one direction, and then a very small, distorted swing in the other direction. A single Class B amplifier is little use as an audiofrequency amplifier, but a pair of valves, sharing the amplification of the signal so that one does one half and the other does the other half, can give a bigger output signal than a solitary Class A amplifier. The grid of a Class B amplifier is Someone looking into a plane mirror sees an undistorted image. The mirror's surface 1s. smooth and flat, so it reflects evenly. So the triode valve 'reflects' the signal wave evenly when it 1s operated in the straight portion of its characteristic.

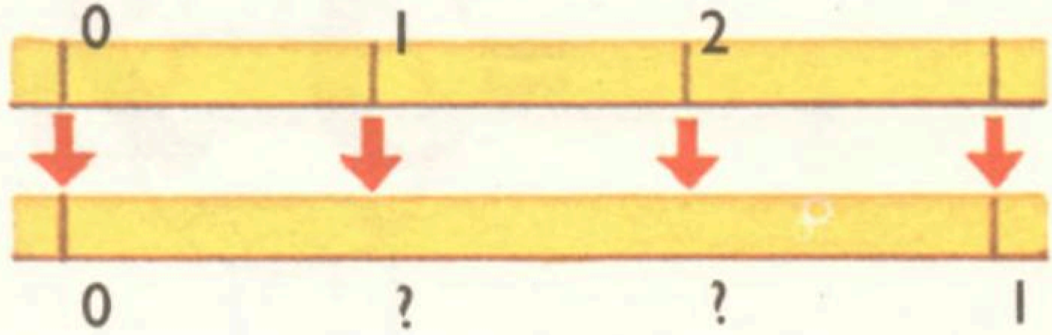

AMPLIFIED
OUTPUT –
UNDISTORTED
INPUT



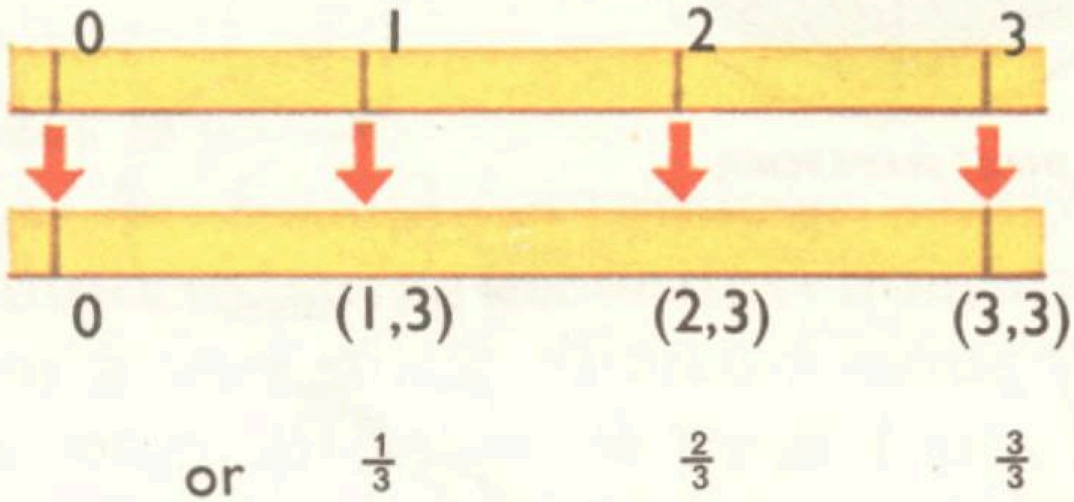


OUTPUT — CLASS B WAVE INPUT WAVE-FORM OUTPUT CLASS B WAVE THESE CIRCUIT DIAGRAMS INCLUDE ONLY THE COMPONENTS NECESSARY ON THE COMPONENTS ARE MISSING The reflection from this curved mirror is distorted. When a valve is operated on a curved part of its characteristic, the incoming wave 1s similarly distorted. This distortion is avoided in a Class A amplifier and used in a Class B push|pull. fhe \§ ; f § \ Hl \ Each triode is biased at the cut-off voltage, so each can ampltfy only half the incoming signal. The signal is divided into two, amplified, and then joined again at the transformer. The result is a Class B pushpull amplifier. One valve deals with current pushes, and the other deals with current pulls. '@ biassed at its ‘cutoff voltage. At this voltage the grid cuts off the current flow from cathode to anode. Although the anode is at a positive voltage, relative to the cathode, and is attracting electrons away from the cathode, the electrons can be repelled back by a negative voltage on the grid. At cutoff, no current flows. When the signal makes the grid even more negative than its cutoff voltage, no current can flow through the valve at all. The anode voltage does not change and this part of the signal does not appear at the valve output. When the signal added to the bias voltage makes the grid less negative, some current can flow. The valve is amplifying again, but can amplify only the positive parts of the signal. But it is possible to connect a ‘matched’ pair of valves together so that each deals with half of the signal. One deals with the current pushes: the other with the current pulls. The result is a Class B ‘Push-Pull amplifier’. Final stages of amplification in a radio set are often done by a Class B pushpull pair of valves (or transistors). They can give a bigger undistorted output ‘swing’. Class C The Grid of a Class C amplifier is biassed even more negatively than it is in a Class B amplifier. The output is a series of short, sharp, highly distorted pulses. Nevertheless is can be used to amplify higher frequency signals (signals of radio frequency). The output signal is so highly distorted that, if it were analyzed, it would be found to contain components of 2, 3, 4 or more @ @ A Class A amplifier. The incoming signal has a sine-wave form. The outgoing signal also has a sine-wave form. A Class B amplifier. The grid is more negative — it is biased just at the grid cut-off voltage. Half of the sine wave is practically undistorted. The other half does not appear at all in the output. A Class C amplifier. The grid ts even more negatively biased. The output is a series of distorted bursts, containing harmonics of the original sine-wave. times the frequency of the input signal. These, the harmonics of the fundamental frequency, can easily be dispensed with in tuned circuits. Grid Current The grid is always biassed negative with respect to the cathode of the valve. The grid should stay negative for the entire voltage swing. This is another limitation on the operating conditions of a valve. Should the grid become positive, it would attract electrons towards it. Electrons would leave the cathode-anode flow, and leave the valve through the grid as a grid current. Grid current represents an unrecoverable loss, and a drain of energy away from the valve. It is also a source of additional distortion, since the partially depleted anode current would not be a true reproduction of the input signal. 2035

MATHEMATICS INTRODUCING the RATIONALS HERE are many other numbers besides the natural numbers (0,1, 2, 3, ...) and the directed numbers, or integers, (which can have negative values). The missing numbers lie in the gaps left between the natural numbers (or integers). They are called the rationals, because thev are defined as ratios (one number divided by another number). The missing numbers can easily be seen by plotting sets of numbers along the number line. The method also involves mapping one set of numbers on to the set of integers on the number line. ings do not correspond to any of the numbers on the yard line. It is obvious that they correspond to a third of a yard. But 4 is neither a natural number nor an integer. What kind of number is it? It is often known as a fraction and, if 1 is divided by 3, by the decimal number 0 *3 (nought - point - three - recurring). Yet another way of writing the number is as an ordered pair of numbers (1,3). These are all ways of writing rational numbers. The ordered pair notation is perhaps the most useful and versatile. Fractions need two numbers (the numeraThe number line for feet. 0 | 2 3 4 5 THE ARROWS MEAN "CORRESPONDS TO' 6 7 8 9 10 OTT 12 The number line for yards. A number of feet correspond to a number of yards. Three feet correspond to one yard, and so on. TY , ; Mapping from one number line to | another. SET OF FEET What happens to I, 2, 4, 5, 7, 8, 10, II, on the number line for feet? Can they © be mapped onto the yards number line? — They certainly do not correspond to i integers (whole numbers) on the yards number line, J Suppose the upper set represents a distance in feet and the lower set a distance in yards. There are three 'foot' markings to every 'yard' marking. So two out of three foot mark2036 tor and denominator of the fraction) to define them. In the usual fractional form, the numbers are set one over the other. In the ordered pair notation, they are set one after the other. Sets of Ordered Pairs The set of rational numbers A small part of the number line is examined on a larger scale. The | foot and 2 feet markings on one number line have no corresponding' markings on the other number line. + (1,3) (2,3) + $ 4 (3,3) Now the lower number line is marked in rational numbers or ordered pairs of numbers. | foot corresponds to $ yard. + is written down as the ordered pair (1,3). The rationals fill in some of the space between the whole number markings. have one thing in common. They are all numbers of thirds. 3 is the denominator of each of the fractions. In other words, they all belong to the set of thirds. In the ordered pairs notation, the same set is written as (1,3), (2-3)s (3.3)s (453)s +> Ordered pairs form a useful set when the second number of the pair is the same for each ordered pair. In particular, this simplifies the rules for adding and subtracting the pairs.



A small part of the number line is examined on a larger scale. The 1 foot and 2 feet markings on one number line have no corresponding markings on the other number line.



Now the lower number line is marked in *rational* numbers or *ordered pairs* of numbers. 1 foot corresponds to $\frac{1}{3}$ yard.

$\frac{1}{3}$ is written down as the ordered pair (1,3). The rationals fill in some of the space between the whole number markings.

Three Ways of Looking at Rationals

As Fractions

$$\tfrac{1}{3}, \tfrac{2}{3}, \tfrac{3}{3}, \tfrac{4}{3}, \ldots$$

The number on top is the *numerator* and the number underneath, the *denominator*.

As Decimals

$$0{\cdot}\dot{3},\ 0{\cdot}\dot{6},\ 1{\cdot}0,\ 1{\cdot}3\dot{3}\ \ldots$$

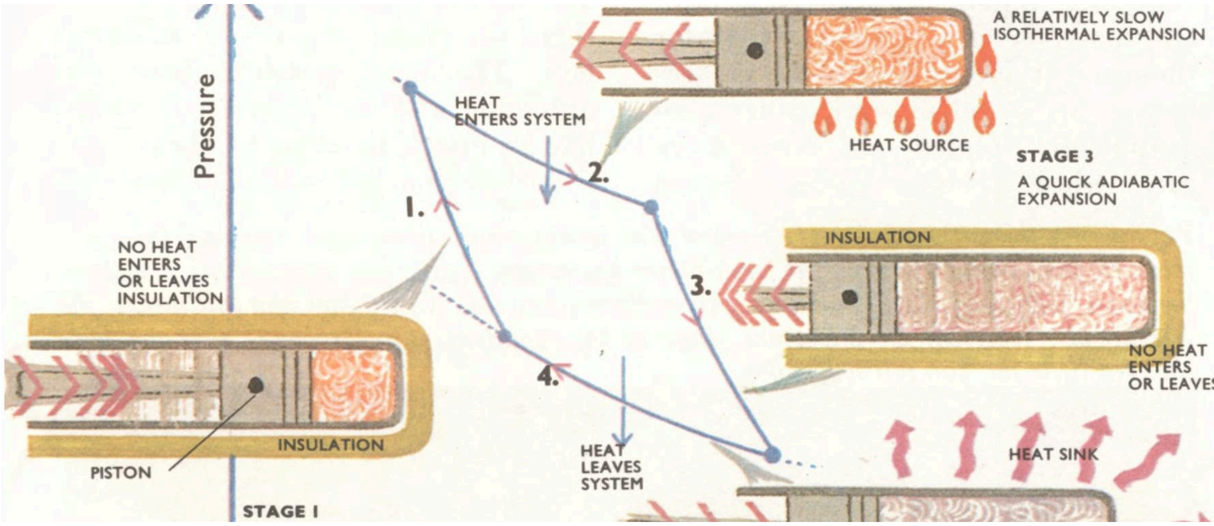
These are the fractions based on tenths, hundredths, thousandths, and so on.

As Ordered Pairs

$$(1, 3), (2, 3), (3, 3), (4, 3), \ldots$$

The first number of the pair corresponds to the numerator of the fraction and the second number, the denominator. It also denotes the set of ordered pairs to which the pair belongs. If the denominator is 3, then the pair belongs to the set of thirds.

FAMOUS SCIENTISTS CARNOT GICLEE IN 1824 a brilliant young French engineer published a paper which has been called 'one of the greatest things in science'. The name of the engineer was Nicholas Léonard Sadi Carnot, and the subject of his paper was the conversion of heat energy into mechanical energy. Carnot had designed an idealized heat engine. The Carnot engine proved useful for showing how the steam engine (which was being developed at the time) can operate most efficiently. It was also a forerunner of the petrol engine and diesel engine. The important point about Carnot's engine was that it operated in a reversible cycle. A gas, called the working substance, absorbs heat, then expands suddenly, rejects some of its heat, and is compressed ready to absorb heat again. If the Carnot engine is put in reverse, all the operations take place exactly in reverse. Compression becomes expansion, absorption becomes rejection, and so on. This is what is meant by reversibility, and all the stages in Carnot's four-stage cycle are completely reversible. The Carnot cycle is and his the most efficient cycle possible at any given temperature. The working substance does work on a piston during the expansion stage. An engine operating on a Carnot cycle gets the maximum amount of work (and so the maximum amount of mechanical energy) out of the hot gas. The sudden expansion of the working substance is an adiabatic expansion. It takes place so suddenly that the working substance is unable to exchange heat with its surroundings. After the working substance has done its work it stays at the temperature of the cooler part of the engine, getting rid of some of its heat, and contracting a little. This part of the process is isothermal (meaning at the same temperature). A sudden adiabatic contraction follows, and then another isothermal stage where the working substance absorbs heat from the hotter part of the engine. Practical heat engine cycles differ considerably from the highly idealized — Carnot cycle. But, theoretically, none can be more efficient than the Carnot engine. They can be just as efficient Carnot's famous cycle. Two isothermal (same-temperature) processes and two quick, adiabatic processes make up the cycle. Any kind of substance can be taken around a Carnot cycle. Pressure NO HEAT ENTERS OR LEAVES A QUICK (ADIABATIC) COMPRESSION STAGE 2 A RELATIVELY SLOW ISOTHERMAL EXPANSION HEAT SOURCE STAGE 3 A QUICK ADIABATIC EXPANSION A SLOW COMPRESSION. HEAT FLOWS OUT TO HEAT SINK Sadi Carnot (1796-1832) investigated the efficiency of cyclic processes. if the stages in the cycle are reversible. This is an important corollory introduced in Carnot's paper. Another is that it does not matter what is used as the working substance — the conditions for maximum efficiency are independant of the working substance. Some of Carnot's ideas were not completely understood, because he confused his terms, notably 'amount of heat', 'the motive power of heat' and a new term which is now known as 'degree of disorder' or 'entropy'. Possibly Carnot's work had been misinterpreted. But Lord Kelvin, the English physicist, saw the importance of the ideas, and rephrased them in their present form. Kelvin was even able to introduce an entirely new concept of temperature, and produced the Kelvin temperature scale, recognized as the only fundamental scale of temperature. The Kelvin temperature scale is based on the efficiency of a Carnot engine operating between two particular temperatures (one where it absorbs heat, one where it rejects it). Carnot was born in Paris in 1796. He studied at the Ecole Polytechnique and became an officer in the Engineers. Unfortunately he caught cholera and died in Paris in 1832, when he was only 36. 2037



A RELATIVELY SLOW
ISOTHERMAL EXPANSION
Pressure
HEAT
ENTERS SYSTEM
HEAT SOURCE
STAGE 3
A QUICK ADIABATIC
EXPANSION
1.
2.
3.
4.
NO HEAT
ENTERS
OR LEAVES
INSULATION
INSULATION
NO HEAT
ENTERS
OR LEAVES
INSULATION
PISTON
HEAT
LEAVES
SYSTEM
HEAT SINK
STAGE 1

Sadi Carnot (1796–1832) investigated the eS RHYNCHOCEPHALIANS TUATARA | DINOSAURS AGO TUATARA OR SPHENODON LIZARD-LIKE BUT FAR REMOVED FROM TRUE-LIZARDS; THESE CREATURES GROW UP TO TWO FEET IN LENGTH 'The tuatara (the maort word means 'spiney back') is the sole survivor of a group older than the dinosaurs. Isolated in New Kealand, tuatara has remained almost unchanged for 200 million years. The crocodiles and turtles may also be regarded as living fossils for they have altered very little from their ancestral forms of Triassic times. The squamata in contrast have changed a great deal. HREE alternatives face every animal or plant species living today. They may, by the process of evolution, change slowly, almost imperceptibly into new forms of life; on the other hand, they may, in the face of competition languish and become extinct; or again, they may remain practically unchanged for hundreds of millions of years. These three alternatives have faced every group of organisms that has ever lived on Earth. Extinction is the usual fate. For instance, of all living land vertebrates 150 million years ago perhaps 1% have descendants existing today. When it comes to counting up living creatures that have actually survived almost unchanged over any great length of time the number remains very small indeed. The process of evolution by natural selection is a ruthless one: Animals which have long flourished may suddenly find themselves in competition with a new kind of creature far bette adapted to the surroundings. Alter-. natively there may be some other fatal 2038 change in the environment — the climate may alter or the staple food may disappear. For animals to exist in much the same way as they have done for millions of years—for them to become living fossils — certain conditions have to be fulfilled. One possibility is that they are ideally adapted to a certain type of surrounding, and that surrounding remains almost constant. Alternatively it is essential that for them to survive, they become separated on an island or some other geographically isolated spot. Away from areas where perhaps better adapted 70 MIELION WEARS | BIOLOGY creatures are appearing, they can continue to exist, if not flourish. Some Invertebrate Living Fossils Some 400 million years ago, buried in muddy deposits about the ancient Ordovician shorelines lived. a small horny shelled brachiopod (lamp-shell) called Lingula. The creature was only an inch or so long and very simple in its structure. Yet nevertheless, almost identical creatures are found today, still burrowing in mud, off the shores of Japan, the Indo-

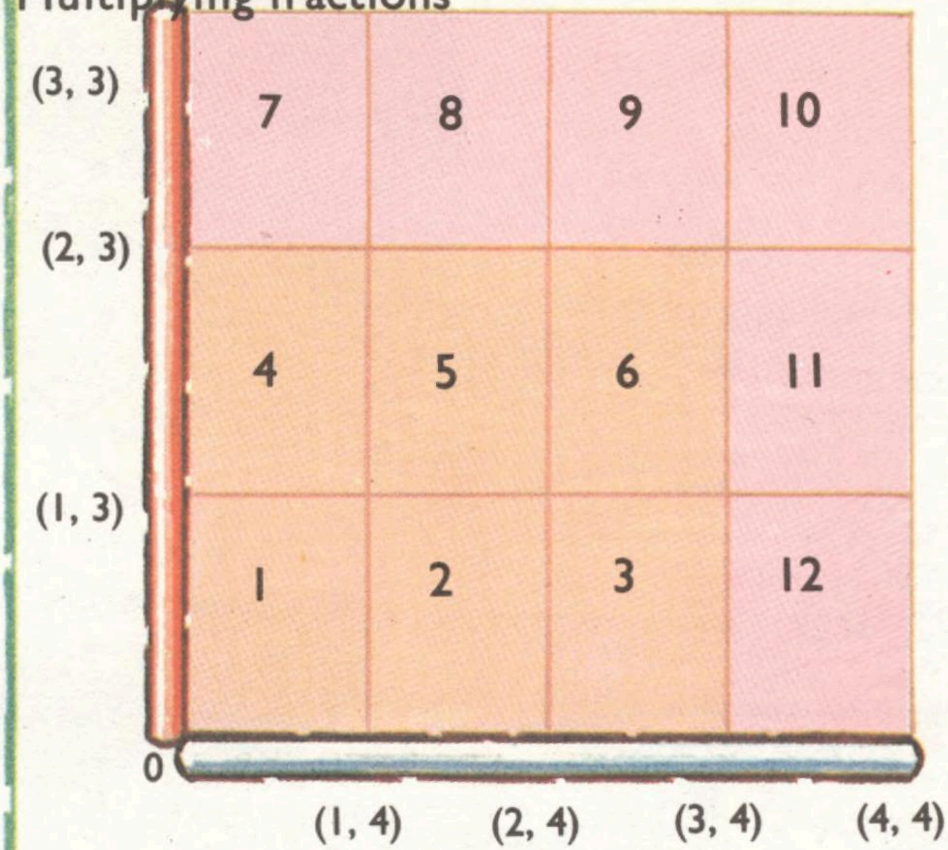
Pacific Islands and Australia. Through hundreds of millions of years, Lingula has remained almost unchanged. Recent investigation of the modern Lingula has shown at least some of the reasons for its success. Buried in muddy shores, coming to the surface only to feed, these creatures are well protected from enemies. But further, the working of their bodies — their physiology — is very well suited to their surroundings. For instance, they can tolerate extremely low oxygen concentration — a condition which would be fatal to many creatures. No offshoots have been so well adapted as the original parent stock, and hence the parent stock has persisted. A very important living fossil was discovered in the West Indies:in 1826. This was a slug-like creature subsequently called Peripatus from its habit of wandering about (Greek, peripatos, a walk around). Other, closely related species have since been found elsewhere in the world. Investigations soon showed Peripatus not to be a slug. The body has many segments, and there are about 20 pairs of unjointed legs. The head possesses jaws and antennae and the creature breathes like an insect, by using trachea. Although a soft-bodied animal, fosPeripatus, this strange worm-arthropod was living 500 million years ago and so was a creature similar to the present-day fresh-water crustacean, Apus. One living fossil seen about the house 1s the stlvertail (probably 350 million years old). Centipedes and millipedes, the Mpyriapods, represent an impo chang ant stage in the evolution of arthropods; they too have

LIMULUS — THE KING CRAB TRILOBITE LARVA OF LIMULUS Two hundred million years ago Limulus the King Crab was found throughout the world. Today, though almost unchanged, it is confined to a few tropical regions. Limulus is an arthropod but its relationship with other arthropods is debatable. Perhaps significantly, its larva is very stmilar to the appearance of extinct trilobites. sils of Pertpatus are known from Silurian rocks 320 million years old. In fact, remains of an animal closely resembling Peripatus go back to the Cambrian period more than 500 million years ago. Zoologically, Peripatus is very important for it does seem to represent a group of animals intermediate between the Annelid worms and the Anthropods, the group which includes the insects. Just why Peripatus has survived unaltered for so long is rather difficult to answer. Rather like wood-lice, it is in continual danger of dessication and survives only in damp conditions; much of the daylight is spent lurking under stones and logs. It is carnivorous by habit, feeding off a variety of small creatures. Perhaps a significant feature in its survival is the degree of parental protection shown — nearly all species of Peripatus bear their young alive. Another 'immortal' which like Peripatus is also a 'missing link', was dredged up in 1956 from a depth of 12,000 feet off the Mexican coast. This animal, called Neopilina is a very primitive mollusc and still bears many Annelid-like characters. Neopilina must represent a group of animals at least 500 million years old. Two other molluscan living fossils are Pleurotomaria, a sea snail known to be living 350 million years ago, and the pearly-shelled cephalopod, Nautilus. Some Vertebrate Living Fossils The continent of Australia was probably separated from other land masses about 150 million years ago. At the time the only mammals living there were the pouched varieties — the marsupials and the egg-laying mammals the monotremes. The more advainced placental mammals developed later, outside of Australia. Consequently for millions of years the monotremes and marsupials existed there, even flourished, when elsewhere, members of their stock were disappearing in the face of competition from the placental mammals. It may not be accurate to say that individually the marsupials are all living fossils; some may have evolved only recently. But nevertheless they are all survivors of a very ancient stock which diverged early in mammalian history. South America was another stronghold of marsupial life but 10 million years ago it became connected with North America. Placental mammals invaded and most of the marsupial population has become extinct. To the west of Australia the island of New Zealand probably became LATIMERIA TODAY'S ONLY KNOWN OCAN Above, the coelacanth; right, one of _ the three surviving species of lungfish. The chart shows the close relationship of coelacanths and lungfishes with the early amphibians. Today's coelacanth closely resembles Mesozoic forms. The Australian lungfish has changed little from Triassiw times. separated from other land masses earlier still. New Zealand does not possess marsupial animals but living on a number of islands off the mainland is an even older creature the Tuatara or Sphenodon. Superficially lizard-like, tuatara belongs in fact to the Rhyncocephaltans, a group of reptiles older than the dinosaurs and possessing many primitive structures. The major factor responsible for the tuatara's survival is the lack of competition it has received from more efficient animals. Other contributory factors may be the long individual life span— 100 years and more —and_ sluggish energy-conser— ving habits. One of the rarest of New Zealand's animals — perhaps the rarest in the world — is the Stephens Island frog. In many respects it resembles other frogs but its feet are not webbed, it has free ribs (unlike other frogs) and its internal blood supplies seem more related to a newt's. The tadpole stage is probably passed within the egg. Such frogs seem direct descendants of the tailed ancestors which resembled present-day salamanders and newts. NOT TO SCALE. 2039

. FRACTIONS and DECIMALS FRACTIONS and decimals often occur in problems. Rarely is the solution a ‘whole number’, a natural number, or an integer: The rules for manipulating fractions and decimals are different from the rules for natural numbers and integers. Adding Fractions and Decimals Add ; to : and the answer is . This can easily be seen by doing the addition along a number line. To add these two fractions, the rule is to add the top half of each fraction to get the top half of the total (the numerator) but leave the lower parts, the denominators, alone. This particular addition is made easy because both fractions belong to the same set of fractions: | EA ee | Paaa | Another way of writing this set of numbers is as ordered pairs of numbers. ‘A is written as the ordered pair (1, 3). All fractions can be written as ordered pairs, the numerator first, the denominator second. So the above set becomes: (1, 3)s (25 3)s (3s 3)s (45 3)5-- 3 The rule for adding ordered pairs is: leave the second number of the pairs alone, but add the first ones. This produces exactly the same result as adding the number pairs in their fraction form. But this method does not work for fractions like 5 and 3 can be written as the ordered pair (1, 4) and 1 as the ordered pair (1, 3). Ordered pairs with 4 as their second number belong to a different set from pairs with 3 as their second number. Rules for adding ordered pairs apply only when both the numbers to be added belong to the same set. The total then belongs to the same set. It is possible to get round this difficulty by writing (1, 4) as (3, 12) and (1, 3) as (4, 12). (3, 12) and (4, 12) both belong to the same set. This works for simple fractions, but not for more complicated ones. It is better to convert the fraction or ordered pair to a decimal fraction. + in deci3 mals is 0°3 (nought-point-three-recurring). If all the numbers are written as decimals, they all belong to the same set, and it is possible to add them easily. The total also belongs to the set of decimal fractions. Multiplying and Dividing Decimals are best for adding and subtracting when the fractions do not belong to the same set. Multiplication and division are, however, best carried out with fractions. No-one would think of multiplying ; by - by turning them 2040 Addition is easy — both belong to the same °o (1, 3) + (2, 3) = (3, 3) = (I, 1) Adding fractions (rational numbers) is easy when both belong tothe same set. ordered pairs (1,3) € 2? with3 their second number ordered pairs ; (2,3)E } with 3 their second number Adding Fractions att) 4) = ———— ——=—_, (1, 3) (2, 3) (3, 3) d,3+0,4=2 The rules for adding fractions (or ordered pairs) do not hold when they belong to different sets. ordered pairs with 4 their second number ordered pairs (1,3) . ung with 3 their second number. Thesymbol ¢ means ‘does not belong to’. Adding Decimals 1, 3) =0-3 Vee (0-3333— — -) 0-3+ 0-25 =0-583 When fractions are turned into decimals it is just like adding in tens. The rules are the same.

into their decimal form 0-25 and 0°3, because it would involve a cumbersome long multiplication. The answer to any fractional multiplication can be seen by putting two number lines at right angles. If 7 is to be multiplied by - ; is marked along one number line and : along the other. + x + is the rectangle bounded by the marked-off parts of the number line. This rectangle is a twelfth of the unit area. If, instead, : is marked off along one number line, and = along the other, the marked-off area is & the unit area. The rule for multiplication is: multiply the two numerators together and the two denominators together to get the final fraction. The rule works for any number of fractions multiplied together. The fraction 2 can then be cancelled to its simplest form, m 2 If the fractions are written as ordered pairs (for example (3, 4) and (2, 3)) the rule is very similar. Multiply the first parts of the pair together, and the second parts of the pair together, to give the first and second parts of the resulting ordered pair. This rule works for the entire set of rational numbers (numbers which can be written as ordered pairs). Division is also best done in fractions. It is the inverse of multiplication. The example for multiplication with rational numbers was: I I I = x< -=— lh 4-3 12 This is very similar to: £ociad 4 12 3 and in fact the two are equivalent statements. +. divided by + 12 gives the same answer as 'what must a third be multiplied by to give the answer as one quarter ?': > ixixsetx3 CC Tene ei (by multiplying both sides of the equation by 3) 5 t a woe 4 (because , and cancel out on the left hand side of the equation) : 4 12° 3 All these equations are equivalent statements. So, when dividing by a fraction the rule is to turn the divisor upside down, and then multiply as ordinary fractions. If the fractions are in the form of ordered pairs (for example (1, 12) and (1, 3)), then the rule is: multiply the first number of the first pair by the second number of the second pair. Then multiply the second of the first pair by the first of the second. This gives the final ordered pair (IX 3,12 X I) = (3, 12) = (1, 4). (I, 4) (2, 4) (3, 4) (4, 4) The unit markings on the number lines enclose unit area. The area enclosed by (2, 3) and (3, 4) is equal to the product of (2, 3) and (3, 4). The third and quarter markings on the number line divide the unit area into twelfths. The (2, 3) and (3, 4) markings enclose 6 of the twelfths, so: 2 x 3 = 6 = 1 3°74 612-2 (2, 3) x (3, 4) = (6, 12) = (I, 2) WZ 2x3 3x4 The rule for multiplying ordered pairs — multiply the first numbers of the pairs, then the second numbers of the pairs. This rule applies even when the fractions do not belong to the same set of ordered pairs. Division is the inverse of multiplication. It is like starting off knowing an area and oneside, and then finding the other side. UNKNOWN pda: | For example: a7 or (I, 2) + (1,3) writing it in ordered pairs. The rule in fraction is: Turn The rule in ordered pairs is: Multiply first number by second and second by first. skied lie Note that 37> zis not the | same as 73 the divisor 5 upside down 2041

Multiplying fractions



The unit markings on the number lines enclose unit area. The area enclosed by (2, 3) and (3, 4) is equal to the product of (2, 3) and (3, 4). The third and quarter markings on the number line divide the unit area into twelfths. The (2, 3) and (3, 4) markings enclose 6 of the twelfths, so:

$$\frac{2}{3} \times \frac{3}{4} = \frac{6}{12} = \frac{1}{2}$$

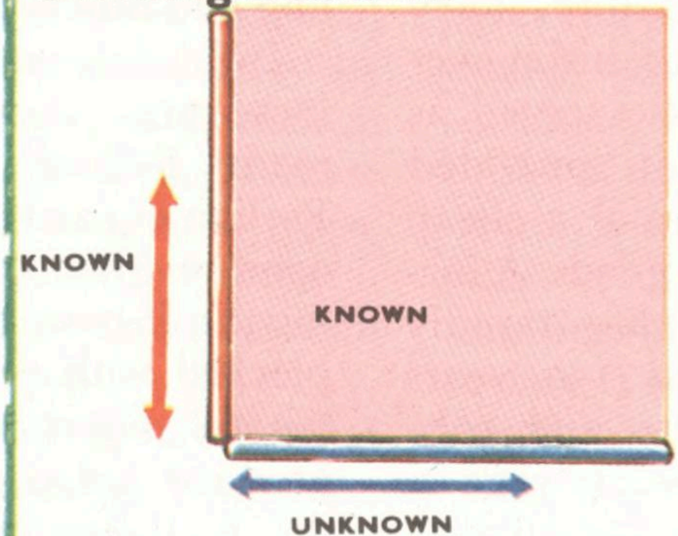
$$(2, 3) \times (3, 4) = (6, 12) = (1, 2)$$

$$2 \times 3 \quad 3 \times 4$$

The rule for multiplying ordered pairs – multiply the first numbers of the pairs, then the second numbers of the pairs.

This rule applies even when the fractions do not belong to the same set of ordered pairs.

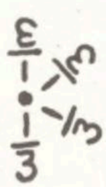
Dividing fractions



Division is the inverse of multiplication. It is like starting off knowing an area and one side, and then finding the other side.

For example: $\frac{1}{2} \div \frac{1}{3}$.
or (1, 2) $\div$ (1, 3) writing it in ordered pairs.

The rule in fraction is: Turn the divisor $\frac{1}{3}$ upside down

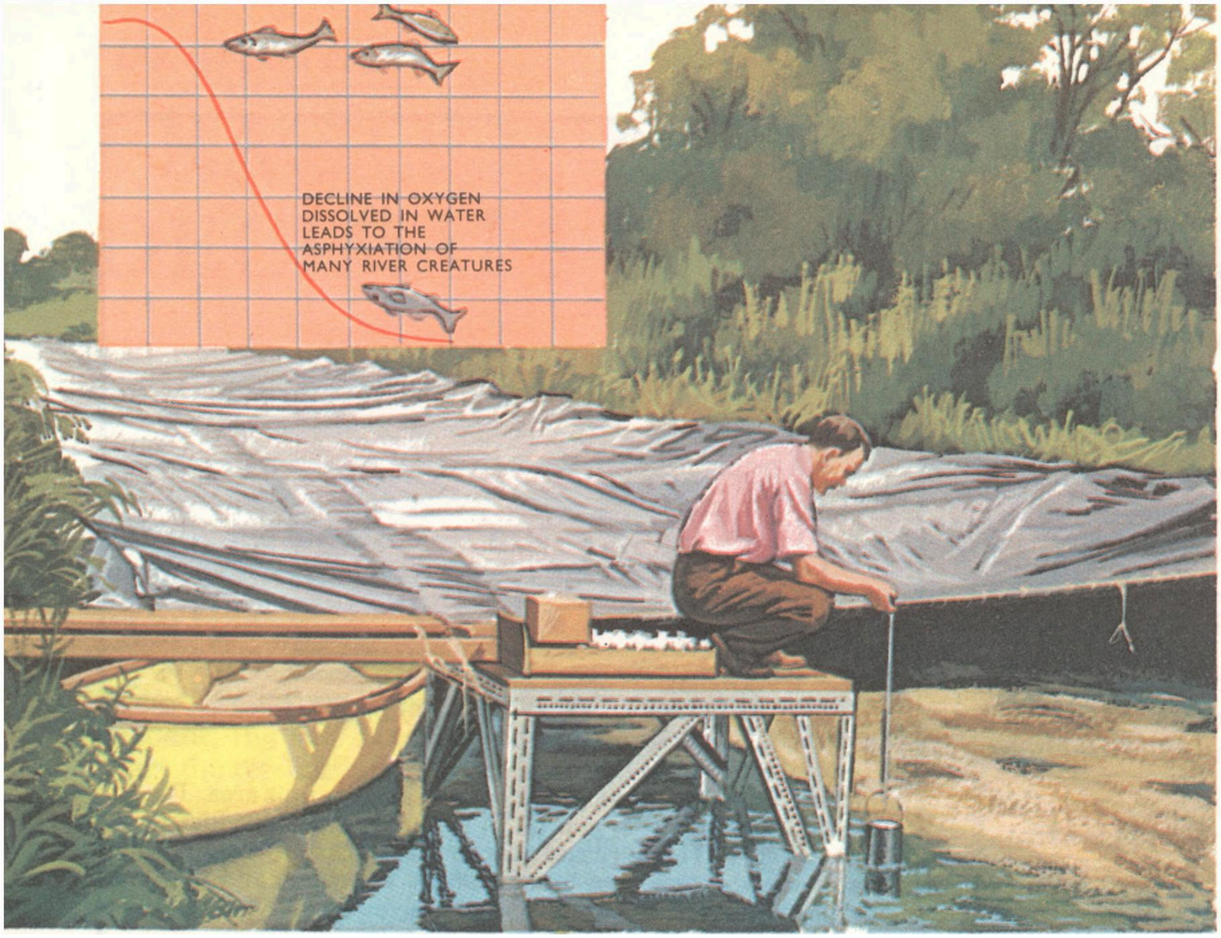


and multiply $\frac{1}{2} \times \frac{3}{1} = \frac{3}{2}$.

The rule in ordered pairs is: Multiply first number by second and second by first.

Note that $\frac{1}{3} \div \frac{1}{2}$ is *not* the same as $\frac{1}{2} \div \frac{1}{3}$.

~"DECLINE-IN-OXYGEN—— DISSOLVED IN WATER LEADS TO THE a yr ASPH a. a MANY RIVER CREATURES GIGANTIC water cycle is in permanent operation. Water evaporates from the sea, forming clouds; then it falls back to the ground again as rain. Over land some of the water is used by plants or stored by Man in reservoirs. But much of it returns directly to the sea as rivers. Rivers have always played an important part in Man's history. They have proved obstacles to his movements, but they have also served as accessible water supplies, provided food in the form of fish and a cheap medium for transporting goods. A further use of rivers that has rapidly expanded over the past 150 years is as a dump for waste and rubbish. For instance, domestic sewage in ever increasing quantities and all sorts of wastes from industries have been piped into them. This continuous polluting of rivers with large quantities of waste has not proved to be entirely to Man's benefit. Rivers soon demonstrate when they have been badly treated. The colour of the water may darken, strong unpleasant smells are given off, certain bacteria increase in number, and dead and dying fish float about at the surface. Conditions are particularly ag2042 Aa seer Ee ie toe wal iy $) < " " % 4 aw WATER POLLUTION gravated in summer time when the quantity of water falls and the amount of polluting matter becomes proportionately greater. Unsightliness is not the only objection to water pollution: public health is endangered; water becomes unsuitable for domestic supplies and agricultural purposes; fish supplies may be annihilated and angling and other recreations threatened. In England as early as the 1850's the increasing quantities of sewage was already fouling the River Thames. Parliament at Westminster could only operate when the windows overlooking the ill-treated Thames waters were blocked with disinfected towels. Distasteful sights and smells certainly produce action, and new laws about the treatment and discharge of sewers were quickly passed; for a time the situation was alleviated. But since those days, signs that all was not well with the Thames and other British rivers have continually occurred. In 1927 the Water Pollution Research Laboratory was set up to investigate and advise on the situation; similar institutes now operate in most countries. Various Public Health Acts and River Board Acts have been passed. These have made it an offence for poisonous APPLIED SCIENCE | The oxygen content of rivers is extremely important. Here an experiment is in progress measuring the relative importance of such factors as animal and plant respiration, and oxidation of decomposing matter. The effect of oxygen produced by plants during photosynthesis ts eliminated for experimental purposes by covering a stretch of river with black plastic sheets. or polluting matter to be discharged into rivers and have given River Boards authority to take samples of affluents for analysis where pollution is suspected. Categories of Polluting Material Investigations by the Water Pollution Research Laboratory are conducted along two lines; firstly, the effluents responsible for pollution are sought and also the quantities in which they become dangerous; secondly, possible remedies are investigated designed to render noxious effluents harmless. The categories of the effluent causing river pollution broadly speaking fall into two natural groups. There is sewage and other domestic refuse which largely consists of organic matter; and there is industrial and commercial wastes largely of an inorganic nature. The organic sewage is piped into rivers by most towns. Only those settlements about the coasts can pipe their wastes directly out to sea. Sewage itselfis usually harmless to animals and plants living in rivers. But it may continue to decompose by bacterial action, and oxygen is needed for the process. The quantity of oxygen in river water, as a result becomes lowered — sometimes with disastrous effects. Fish and other freshwater organisms die by asphyxiation; decomposition of later sewage is only partially completed and hydrogen sulphide and other unpleasant gases are given off. The second category is the array of inorganic wastes discharged from a whole range of industrial plant — gas, chemical and dye works, collieries, iron and steel foundries, food factories, paper mills, oil refineries, etc. Such substances are usually more direct in their activities. Solid particles such as coal dust and cinders kill by blanketing bottom-living organ


DECLINE IN OXYGEN
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MANY RIVER CREATURES

Solid matter discharged from industrial works can blanket the inhabitants of river beds. Chemicals in solution may be poisonous to life. isms and river plants, and clogging the gills of fishes. Chemical compounds such as lead, zinc, copper and ammonium salts have toxic properties and destroy life by upsetting the biochemistry of organisms. Some substances, like sewage, have a deoxygenizing effect. Synthetic detergents for instance lower the rate at which oxygen is absorbed into river waters from the atmosphere. A combination of low oxygen concentration and toxic chemicals is particularly lethal, for the lack of oxygen increases the potency of poisons. In some industrial stretches of river however, pollution by chemicals kills practically 100% of living organisms despite the fact that oxygen content of the water is high. Cures and Preventions Some cures for excessive river pollution immediately present themselves. Sewage must be treated -before its release into rivers, so that little further decomposition can take place. The oxygen content of rivers is not then Scenes ltke this will not be seen much longer. Detergents can now be made from chemicals which will decompose under bacterial action. noticeably lowered. Methods of purifying sewage involve separating out coarse and fine suspended materials using screens and sedimentation beds and then seeing that remaining material is almost completely decomposed by _ bacterial action. When discharged into rivers, sewage should be liquid in its form, and harmless in its properties. The effects of various concentrations of organic and inorganic chemicals on the lives of river organisms needs more detailed study. Research by carefully arranged experiments and accurate observations determine just what substances are potential pollutors. Poisons are not just tested out in large quantities. What happens when minute quantities are present over very long periods is also investigated. Other measurable influencies on the toxicity of poisons are the temperature, oxygen content, and acidity of the river waters themselves. The factors vary with both the locality of the river and the time of year. Whether poisonous industrial wastes are piped into rivers directly, or via sewage works they must first undergo treatment. Noxious substances must be removed or be present only in quantities below the estimated danger level. Techniques vary according to the waste. Soluble lead, zinc, copper and chromium salts are precipitated by the addition of suitable chemical compounds. Acids are neutralized by lime. Catchpits remove solids such as coal dust and waste china clay. Grease and oils are separated out in traps. Tanks also must be provided so that if necessary outward flow into rivers can be immediately stopped. Some sources of pollution are quite accidental. The pesticides, fungicides, 4 . New Looks for Detergents Synthetic detergents are now used for a variety of household as well as commercial purposes. During 1963 about 150,000 tons were used by housewives alone. These detergents pose a problem for the sanitary engineer. They are chemically very stable and are not attacked and decomposed by bacterial action. Consequently, foam impairs sewage plants, mars rivers and most important of all, lowers the rate at which oxygen is absorbed from the air by water. Today the problem, at least chemically, is solved. The most important ingredient of the detergent is the 'sudsing agent', a chemical which operates by lowering the surface tension of the water, enabling it to froth. Unfortunately the common agent used (tetrapropylene benzene sulphonate) is structurally very stable and is not decomposed by bacterial action. Substitutes equally as effective but which can be decomposed by bacteria are now available. Detergents containing these substitutes are likely to be enforced by law in the near future. and insecticides sprayed in large quantities, often indiscriminately, on crops and woodlands eventually find themselves washed into rivers where their lethal properties may still be possessed. Here then is further evidence for enforcing the controlled use of chemicals on our agricultural lands. Legislation often takes a long time to come into force, and damage done over the past decades may take many many years to be eradicated. For instance, today 100 miles of the 550 miles of the River Trent, Nottinghamshire, is so polluted that no life can exist at all. But signs of improvement are beginning to be felt. The Thames below Teddington lock is also still highly polluted. Once a fine salmon river, no salmon has been caught in this stretch of water for over 100 years. But surveys show that at last by applying every sort of remedy, the oxygen content of the water is rising once 'more. Governments are now well aroused to the dangers of water pollution and many Official bodies as well as preservation societies and angling associations keep careful watch over waterways. With the amount of sewage and waste chemicals likely to increase in future years, this is just as well. 2043

New Looks for Detergents

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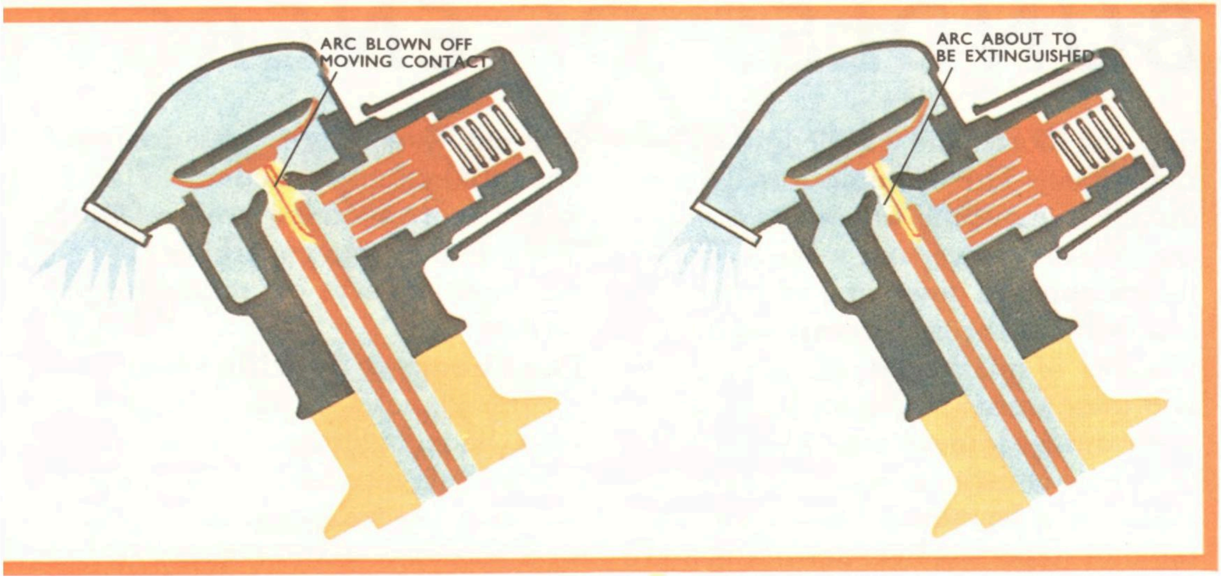
ELECTRICITY VERYONE is familiar with an electrical switch, and there is hardly an item of electrical apparatus which does not contain one kind or another. A switch works simply by moving a piece of conducting metal so that it touches both of a pair of contacts. The electricity can then flow from one to the other. The contacts are mounted on an insulating base, usually plastic or porcelain. A very simple kind of switch is the knife-switch which looks rather like a paper-cutting guillotine. The blade is hinged to one contact and can be moved down to connect with the other. When the blade is raised, the electricity could flow only through MOVING ONTACT FIXED CONTACT A very simple switch, the knife is sometimes used for switching small electric currents on or off. The symbol normally used in electrical circuit diagrams is of this kind of switch. either the base or the air between the two contacts, but neither are good conductors, so the current is switched off. In small switches (e.g. for low voltage lighting systems) there are no great difficulties. But switches controlling great amounts of power in industry and the public supply system need careful design. Firstly, of course, the moving contact, the fixed contacts and terminals and the area of contact between them must be large enough for the current to be carried, otherwise they will over2044 MOVING CONTAGHS Inside the air-blast circuit br and the arc is struck be PISTON TOP OF INSULATOR breaker, the pressure of the air forces back the I en the two specially hardened contacts. The air-blast carries the arc up towards the top of the air passage, until it is extinguished. ARC LENGTHENS ng contacts, SWITCHES and heat and, in extreme cases, melt. The distance between the fixed and moving contacts when open (i.e. when the switch is 'off') must be too large for the current to 'jump across' in the form of a spark or arc. Also the insulating material used in the switch must be a good insulator at high voltages. It must also not allow 'tracking' over its surface especially in adverse (such as very damp) conditions. This ability of electricity to jump gaps produces difficult problems with all switchgear controlling high voltage current, and the difficulty becomes very serious with those switches, for example, that control the flow of electricity through the lines of the national grid. (These can carry enough power for whole towns.) The problem is arcing. No switch can open instantaneously — its moving contact must move at a finite speed, and so for a very brief moment as it opens, the gap between the fixed and moving contacts will be too small to prevent the electricity from jumping across. This it does, and forms an arc. An arc consists of an electric current flowing through air which it has tonized. The air then becomes a conductor of electricity. If the current passing through the switch at the time it opens is large, the arc will also be large, and more stable, and if the current and voltage were sufficiently high, the arc might continue even when the switch was fully open, and it would very soon burn or melt the whole switch. In most switches, the current passing is insufficient to maintain an arc of any size. The contacts are made to fly apart very quickly, and so any arc formed can exist only for an extremely brief time, during which it has little opportunity to do much damage to the contacts (although it is often possible to see the burning caused after a great deal of use). Most domestic switches are made so that, as the operating lever is moved, energy is first stored in a spring and then all used at once to flick the moving contact quickly across. In large switches, such as those used in the switching stations to control the flow of electricity in the Grid, or in the many factories which use enormous amounts of electrical power, or in the supply to electrified railways, some means of switching off the heavy flows of current is essential, particularly in an emergency. In switches of this kind (usually called circuit breakers because they are made to open automatically, to protect the circuit, and the supply, in the same way as a fuse, should anything go wrong such as a short circuit), some means of extinguishing the arc is essential. A very commonly used device is the azr-blast type of circuit breaker. In this, a blast of air is used to blow the arc away. Such circuit breakers are usually installed to work with an isolator, a simple switch working on the principle of the knifeswitch. The air-blast circuit breaker stops the flow of current, and then the

ARC BLOWN OFF NG Ovi c ARC ABOUT TO E EXTINGUIS CIRCUIT BREAKERS isolator swings open. Since no current is flowing there will be no arcing. When the isolator is fully open the circuit breaker contacts close again. Often the isolator is used to re-connect, or re-make the circuit, since there is no arcing problem. In this case it is sometimes called an zso-maker. The air-blast circuit breaker is fairly complicated, and looks very much more so, since up to twelve breaks are connected together in series to deal with high voltage. Each unit is operated by air pressure, supplied from a large reservoir very close by, often forming the base of the equipment. This is necessary so that the air loses little pressure or speed in flowing to and through the breaker. In operation an air valve releases a blast of air at about 350 pounds per square inch pressure through the breaker. This pressure acts on a piston connected to the moving contacts which are forced back and so disconnected from the fixed contacts. The arc is struck between two parts of each contact, deliberately made to project and so part last of all, which are made of specially hardened copper. The airblast is passing between these contacts and carries the arc up towards the top of the apparatus until it is finally extinguished. A fixed time (a fraction of a second) after the air-blast starts the isolator opens. As soon as it is fully open the air valve again closes and the breaker contacts, which were held open only by the pressure of air, close again. Apart from blowing the arc away, the air pressure makes conditions most unfavourable for the formation of an arc, which takes place most easily in very low pressures, but as the pressure rises, the voltage necessary to strike and maintain an arc also rises. Another type of circuit-breaker in quite frequent use is the ozl-immersed type. The contacts are kept submerged in special oil which make formation of an arc extremely difficult. As the contacts open, it is arranged that a jet of oil passes between them, so, as in the air-blast type, carrying the arc away. Circuit breakers can be opened simply by operating a control, which sets the device going (a typical 275,000 volt air-blast type takes about $ second to open fully). But it is not often that the need arises for one to be opened. They are designed to cut off very large supplies of power. Normally this power is controlled by other means, for example, in the public supply by smaller switches in local substations. Opening a main circuit breaker may be necessary in order to protect all the equipment supplying the electricity. Failure so to protect it may destroy it. These circuit breakers are arranged to open automatically in certain circumstances to protect the supply, for speed is essential should a fault occur. One condition which will cause the contact-breaker to trip (i.e. open automatically) is if the current passing through it exceeds a certain amount. They would also open if an overhead power line were cut. Lightning, striking overhead lines, while not usually very damaging, can cause a surge of current which will trip the breakers, and to avoid the circuit being disconnected for an unnecessarily long time from the supply system, in this case the control equipment can be arranged so as to automatically reclose the circuit breaker. If the fault is still present the breaker will then immediately trip again, and this time remain open, the control equipment having been automatically locked out. A circuit breaker for a 275,000 volt main transmission line. There are eight separate atrblast interruptors which all open simultaneously, and then the isolator swings open. AIR RESERVOIR AIR BLAST PASSES UP INSIDE INSULATOR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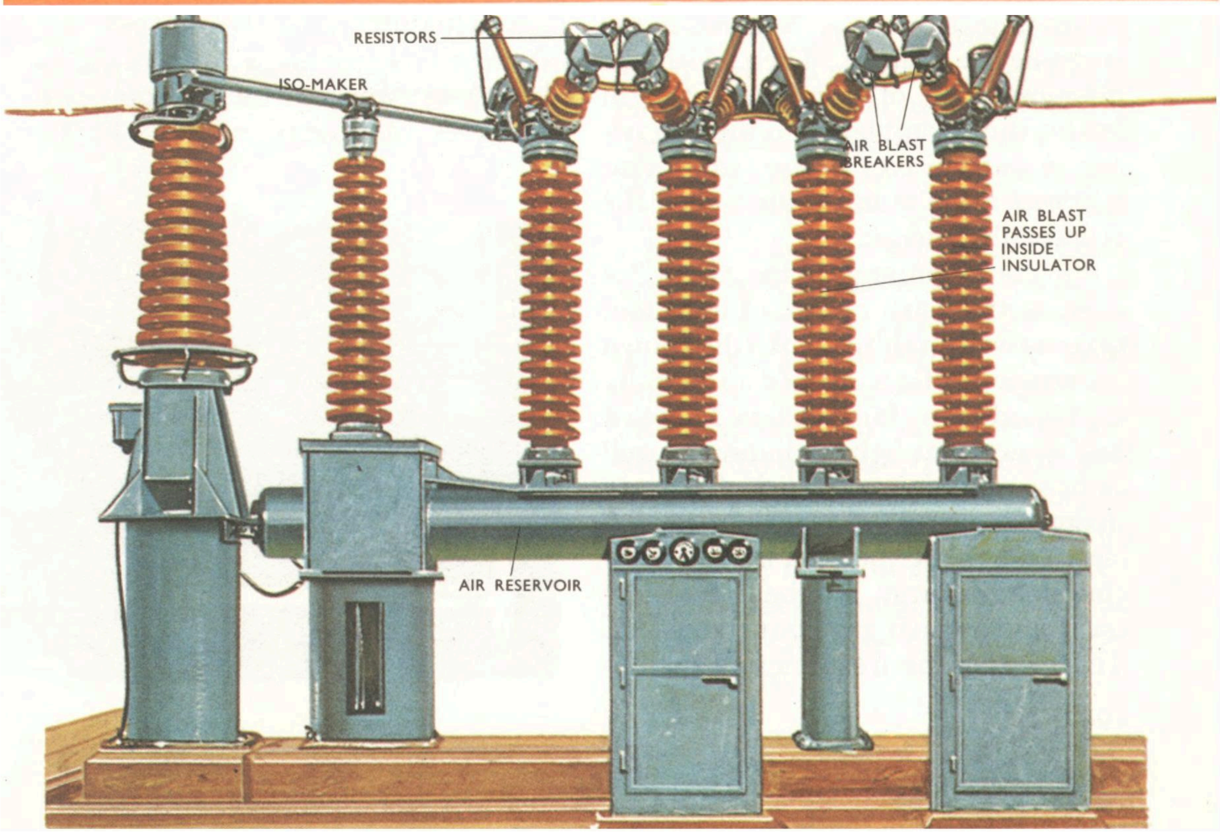
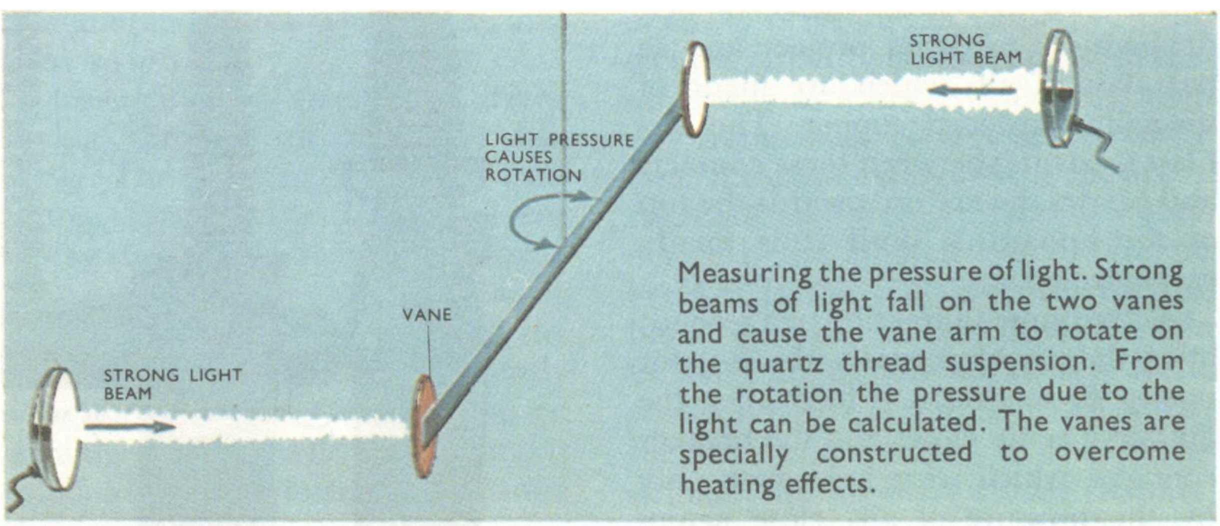

RESISTORS
ISO-MAKER
AIR BLAST
BREAKERS
AIR BLAST
PASSES UP
INSIDE
INSULATOR
AIR RESERVOIR

| ATOMIC PHYSICS | QUANTA — BUNDLES OF ENERGY F proof were ever needed that the scientist must always be openminded and prepared to change his ideas when new facts are discovered, it is to be found in the way that theories of the nature of light have changed through the centuries. First, it was thought that light was thrown out from a luminous body in the form of light corpuscles — tiny round pellets. This was believed by the ancient Greeks and the theory still held sway in the time of Sir Isaac Newton, in the seventeenth century. Then, with the demonstration of interference of light beams the resemblance to the behaviour of water waves was so strong that the wave theory of light superseded the corpuscular theory. By the beginning of the twentieth century the wave nature of light had become well established. All the known experimental facts seemed to be fully explained by it. James Clerk Maxwell had shown that light radiation as well as heat radiation consisted of electromagnetic waves. According to the electromagnetic theory a heated body contained a large number of atomic oscillators, which gave out the waves. The greater the frequency of oscilla2046 tion, the shorter the wavelength of the radiation. The atomic oscillators should give out the waves continuously, like the waves given out when a cork is bobbed up and down in a water tank. Maxwell's theory was expressed in a number of mathematical equations, and these were applied by Lord Rayleigh and Sir James Jeans, who worked out how much energy should be expected to be radiated by a black body radiator (a perfect heat absorbing and emitting cavity) at different wavelengths. One of the predictions of the Rayleigh-Jeans law of radiation was that as the wavelength of radiation was made smaller and smaller, the energy of radiation should get larger. The actual radiation law, found by experiment, was in complete disagreement with this conclusion. There was obviously something wrong with the theory that the law was based on. Then Max Planck in 1900, made a quite revolutionary suggestion — that the radiation was not given out in a continuous train of waves, but in the form of radiation packets called quanta. The atomic oscillators still existed in this theory but they threw out energy in short, sharp, bursts. The energy of each quantum depended on the /frequency of the radiation — the higher the frequency (i.e. at shorter wavelengths) the greater the energy of each quantum. These ideas were expressed in mathematical form by Planck and from his equations a new radiation law

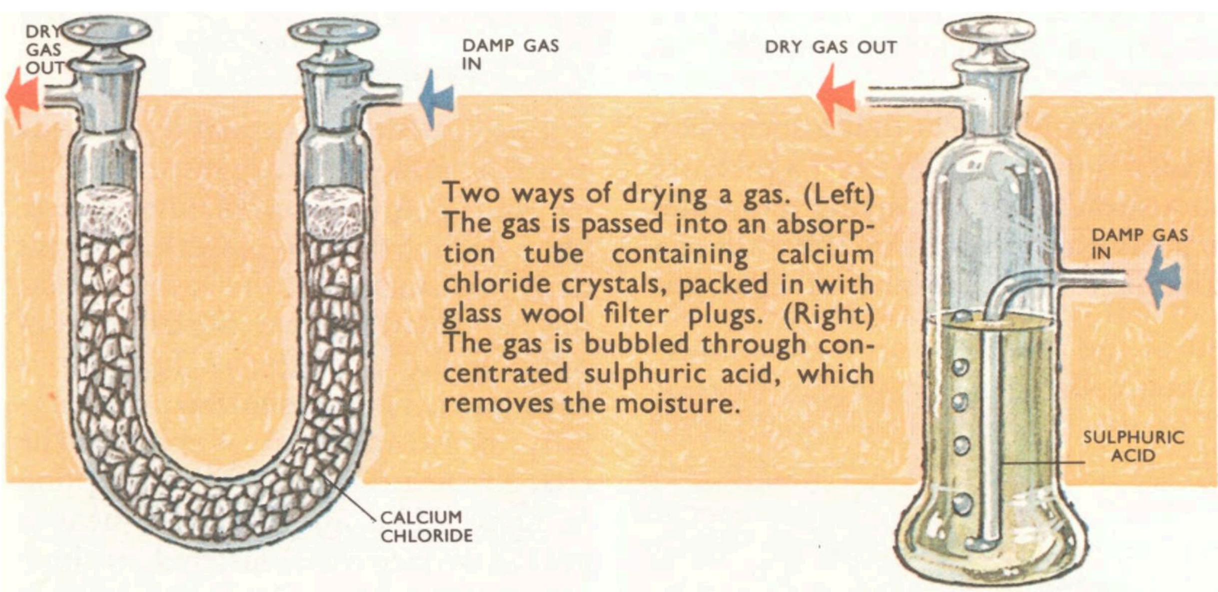
was worked out. From. this, amounts of energy emitted at the different wavelengths was predicted. The predictions, based on Planck's theory, corresponded well with the results found from experiment. So the quantum theory of radiation was verified. The Theory of Specific Heat After Planck's success in explaining black body radiation, his theory was applied to a number of other physical facts that needed clarification. One of these was the variation of specific heat of materials with temperature. Beforehand, it had been assumed that as a mass of material was heated up, the heat energy was absorbed by the solid, and the molecules in the solid took up the energy by vibrating with vibrations of increased amplitude. The amount of heat energy absorbed by a gram of the material for one degree rise in temperature is called the specific heat. If each particle simply vibrates more and more vigorously with a rise in temperature there is no reason to expect that the specific heat would change with a rise in temperature. The heat needed to raise one gram of copper from 19°C to 20°C should be the same as that needed to raise it from 99°C to 100°C. The experimental results of measurements of specific heats showed, in fact, that specific heat did change with a change in temperature, so the simple theory of the continuously vibrating particle could not be correct. Then Albert Einstein suggested that the heat energy possessed by the particles must be in the form, once again, of energy packets — quanta. A



Measuring the pressure of light. Strong beams of light fall on the two vanes and cause the vane arm to rotate on the quartz thread suspension. From the rotation the pressure due to the light can be calculated. The vanes are specially constructed to overcome heating effects.

SOURCE OF BLACK BODY RADIATION ad Finding the change in radiation energy with change in wavelength. The heat radiation is given out by a black body radiator and is split into the different wavelengths by the lens and prism (spectrometer). The thermopile is moved along to pick up the different wavelength radiations and the meter reading indicates the BROK FOUN, METER READINGS energy of the radiation at each position. The energy is plotted against wavelength on a graph. The experimental (full-line) curve is in accordance with Planck's theory — the wave theory would lead to the dotted line curve, not found in . practice. mathematical equation that expressed this idea was worked out, and the change in specific heat with temperature was predicted from this equation. The predicted results were found to correspond to the results found by experiment. Once again the idea that energy exists in the form of quanta was verified. The Photoelectric Effect When light falls on a metal surface electrons can sometimes be dislodged. This is the photoelectric effect. The energy of the dislodged electrons can quite easily be studied in the laboratory. The results of such experiments were studied by Albert Einstein. One result that caught his attention was that even when a very weak beam of light was used, an electron was sometimes emitted. Sometimes this would happen quickly and other times there would be a long interval before the electron was dislodged. Einstein pointed out that this result could only be explained if the light was in the form of quanta, (called photons). If the light were in the form of a wave, the energy would be spread evenly over a large area, and in some of the experiments it would take much longer to dislodge an electron than was actually the case. So once again the quantum theory was able to explain an experimental result that the classical theory could not. The Pressure of Light The application of the quantum theory of energy was very successful in explaining black body radiation, specific heats, and the photoelectric effect. In all of these examples the theory had to be applied through the use of mathematical equations. It is also possible to show, by means of a fairly simple experiment, that light consists of a stream of particle-like bundles of energy. One would expect that a succession of such bundles would exert a pressure on a surface just as moving molecules in a gas exert pressure on the walls of a container. The pressure of light has TEMPERATURE RISE FOUND USING THERMOCOUPLE HEATING COIL The specific heat of a solid at different temperatures can be found by measuring the rise in temperature when a known amount of heat is absorbed by the solid from a heating coil. If the results of such experiments are plotted on a graph the curve shown results. These results can only be explained by the quantum theory of specific heats — the older theories predict a straight horizontal line curve. WAVE THEORY ADS DS TO SAME CURVE CURVE. NOT AS FOUND IN IN EXPERIMEN EXPERIMENT ENERGY OF WAVELENGTHS OF RADIATION been demonstrated in a laboratory experiment. The apparatus used has to be very sensitive because the actual ? I ressure is only ——————— dyne per P bs 1,000,000 bile square centimetre. Two vanes are attached to the ends of a bar, and the bar is suspended, at its mid point, from a fine quartz fibre. Light from two powerful sources strikes the surface of the vanes so that the bar rotates. From the amount of rotation the pressure of the light is calculated. The vane surfaces are carefully constructed to eliminate the effects of heating due to the absorption of light. (The air molecules near a heated surface move faster than those at an unheated surface so they exert extra pressure.) Counting Photons Individual light quanta (photons) may be counted using an electronic counter. In this, a photon strikes a metal wire and releases an electron by photo-emission. The wire is sealed in the chamber along the axis of a metal cylinder, and a voltage difference is maintained between the wire and the cylinder. The released electron passes to the positive cylinder and in doing so releases more electrons as, en route, it ionizes gas molecules. A small avalanche of electrons is formed and this produces a current pulse that can be detected by an electronic circuit. By adjusting the voltage difference between the wire and the cylinder, the cell can be made to give one current pulse for each photon that enters. So, individual photons can be detected and counted. 2047

INORGANIC CHEMISTRY Drying Agents ONE of the most common impurities to be found in newly prepared chemicals is water. Great care is taken to remove the water because the smallest trace might, in an otherwise pure substance, have profound effects on its chemical properties. Chlorine gas, for example, when absolutely pure and dry, is not particularly reactive, but when the slightest trace of moisture is present it becomes one of the most reactive gases known. To remove moisture, the chemist has at his disposal a number of drying agents (desiccants). The choice of a particular agent depends on the parDAMP GAS IN ticular application. One might be more effective than another, but may not be used because it reacts with the chemical it is supposed to be drying. Some drying agents merely take water into their surface structures — silica gel, for example. Others remove water as a result of a chemical reaction. For example, concentrated sulphuric acid will remove water of crystallization from copper sulphate crystals, turning it from the blue form to the white anhydrous form: CuSO4, 5H,O -> CuSO, + 5H,O blue white In an even more powerful reaction, it drags water right out of the sugar molecule, leaving behind a black mass of carbon: Cy2He2O13 —+> 12C + 11H,O sucrose carbon water Though it is obviously a good drying agent sulphuric acid must be used with care because of its properties as a powerful acid and oxidizing agent. The common gases which can be dried by concentrated sulphuric acid include nitrogen, oxygen, carbon dioxide, and hydrogen but others — ammonia, nitrogen dioxide, hydrogen bromide, hydrogen chloride, and hydrogen sulphide — cannot be dried in this way. All of these gases react with the acid. There are a number of solid desiccants that might be used in the laboratory. These are often anhydrous forms of crystals that take up water of crystallisation, although phosphorous pentoxide — a very efficient desiccant, is turned into a wet mass of phosphoric acid by the water it absorbs. A more commonly used one is anhydrous calcium chloride. To dry a gas the solid is placed in a drying tube and the gas is passed over it. To dry liquids (particularly organic compounds) the solid calcium chloride is dropped into the liquid and allowed to stand for a few hours. Then the liquid is passed through a coarse glass wool filter and the solid, which has taken up the water from the organic liquid, is separated out. Other solids used in organic chemistry are anhydrous magnesium sulphate and anhydrous sodium sulphate. These are slower in action than calcium chloride. The most drastic way of drying a solid is to heat it to dryness, but this Desiccants The following is a list of the most commonly-used drying agents: phosphorous pentoxide magnesium sulphate melted caustic potash concentrated sulphuric acid silica gel calcium oxide anhydrous calcium chloride zinc chloride copper sulphate There is an enormous variation in the effectiveness of these drying agents — phosphorous pentoxide leaves behind | in 2 885790000 of the quantity of water that an inefficient drying agent like copper sulphate does. cannot always be done because the solid might decompose at high temperatures. A safer method is to dry the solid in a desiccator. This is a sealed vessel in which the damp solid is placed on a watch glass. The drying agent — sulphuric acid, calcium chloride, or phosphorous pentoxide is placed in the base of the desiccator. The water vapour in the atmosphere in the desiccator is quickly absorbed by the desiccant and as more comes off the damp solid this is removed, too. The desiccator is also used as a cooling vessel. When a solid is dried by heating, it is essential that it cools in a dry atmosphere to prevent absorption of moisture from the atmosphere. This is ensured by allowing it to cool off in the dry atmosphere of the desiccator. The solid to be dried is placed in a sealed dessicator. The atmosphere is kept dry by the dessicant, which may be one of a number of compounds, depending on the degree of dryness required. 2048 DRYING AGENT IN DISH



Two ways of drying a gas. (Left) The gas is passed into an absorption tube containing calcium chloride crystals, packed in with glass wool filter plugs. (Right) The gas is bubbled through concentrated sulphuric acid, which removes the moisture.

| APPLIED SCIENCE | STREAMLINING - Ships and Submarines WHEN a ship travels through water, it experiences two kinds of resistance to its motion. One is the frictional drag of the water around the hull of the ship, and is called the skin resistance. The other major form of resistance arises from wave-making. Although the bow wave of a ship forging ahead at high speed may look very dramatic, it represents an enormous wastage of energy. Energy is expended in overcoming all forms of resistance. Both skin resistance and wave-making depend on the size, the speed and the shape of a ship. The size and speed range of a ship are more or less fixed by the time the ship gets to the drawing-board stage — the naval architect knows whether he is dealing with a slow tramp steamer, a fast frigate, or a large, fast passenger liner. The shape is the remaining variable. A good shape can cut down the resistance losses considerably. The shape of a ship follows basic patterns of smooth streamlined curves, designed to cause as little disturbance as possible as the ship cleaves through water. Other general features of the shape are determined by the speed and function of the ship—a tanker, for example, needs a bulkier, wider hull than a frigate because it requires more storage space. The wide tanker hull is designed for slower maximum speeds than the narrower frigate hull. However, the final details of the streamlining are always decided after tests with models in towing tanks. Although approximate values for resistance and efficiency can be calculated theoretically, tests of models give more realistic answers. There should be no bumps, holes, or sharp discontinuities in the sides of the ship, for smooth flow is possible only if the ship has smooth streamlined sides. The bow is tapered or rounded to part the flow lines of water. As the ship moves, it carries along with it a thin skin, or boundary layer of water. This boundary layer is moving at the same speed as the ship: beyond it is stationary water. The stationary water exerts a frictional drag (a viscous drag) on the moving water. This is the origin of the skin resistance. The stern of the ship is usually rounded or tapered, too, so that the flow lines of water can join together with as little Streamlined flow around a hull. Some frictional drag is inevitable because water is, to some extent, viscous. This resists the motion, and leads to an energy wastage. The boundary layer thickens near the rear end of the hull. HIGH PRESSURE LOWER PRESSURE The other source of energy wastage — wave-making. All surface craft make waves,

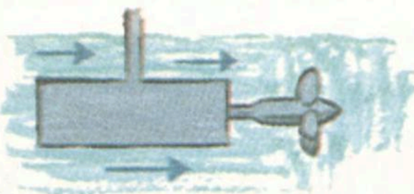
which are mainly the result of increased pressure at bow and stern, reduced pressure midships. Wave-making 1s largely eliminated in submarines. energy-consuming turbulence as possible. Here there are bound to be complications in the streamlining, to allow for the siting of the propellers, and the rudder. In designing a ship the naval architect aims for a hull shape that will produce as little drag as possible, but there are a number of other factors which must be considered eS a a aa ea a a a a Block Coefficients The block coefficient is the ratio of the volume of hull below the water line to the volume of a block with the hull's maximum length, width and depth. FRIGATE BLOCK COEFFICIENT BOUT 0-6. fish SPEED — OW BLOCK COEFFICIENT CURVED SIDES JUST OVER HALF OF BLOCK FILLED BLOCK OIL-TANKER BLOCK COEFFICIENT ABOUT 0-8, LOWER SPEED, ~ HIGHER BLOCK COEFFICIENT HIGH STORAGE CAPACITY ALMOST STRAIGHT SIDES, FILLS NEARLY ALL * BLOCK HIGH PRESSURE too, and which often outweigh the importance of streamlining. A hull that tapers gently from fore to midships, and then back from midships to stern, will give a low drag, and it will also possess a low block coefficient. This is the ratio of the actual underwater volume of the ship compared to the volume of a block of the same length, with a depth equal to the draught of the vessel, and a width equal to the maximum width of the vessel. The draught of the ship is taken as the depth below the Load Water Level. Oil tankers have high block coefficients. They are little more than boxes, with only a small percentage of the total length taken up in the curving of the bow. A typical oil tanker 'occupies' around eight-tenths of the block, so its block coefficient is about 0-8. Military vessels have comparatively low block coefficients because their emphasis is on speed and range, not on carrying capacity. The block coefficient of a frigate is about 0-6. Nearly all its length is curved. Similar sleek, graceful lines are also to be found in passenger liners and in cargo ships which transport perishable goods. Again, speed is of paramount importance, so streamlining takes precedence over high block coefficients. The skin resistance increases rapidly as the speed increases — it is approximately proportional to the cube of the speed. Even a small increase in speed must be paid for with a comparatively large increase in the energy needed to overcome the resistance. Wave Making and Bow Forms The pressure is high at the bottom of the fore-end of the hull, and again high at the bottom of the rear end. This pressure disturbance leads to the creation of waves on the water-surface. Like the skin resistance, the wavemaking depends on the shape (the bottom of the fore-end of the hull is especially important), on the speed. of the ship, and on its length. Wave-making is proportional to speed (or velocity). The higher the speed, the bigger the energy wastage through wave-making. But the longer the ship, the smaller the loss. Roughly speaking, wave-making is inversely proportional 'to the square root of the length. The quantity # (velocity divided by the square root of the length) is a very useful one to the ship designer. When he talks of a 'fast' ship he does not necessarily mean one which travels at high speed, but rather one for which the quantity fis high. Some ships are being designed with bulbous swellings below the water line at the bow end (bulbous bows, or ram bows), which can reduce the energy wastage due to wavemaking by creating a wave disturbance which interferes The hull is towed along to determine its drag-skin resistance and ~wave-making. This is carried out in perfectly still water. The propulsion system is added and the model moves under its —own power. The overall efficiency of propulsion can now be deduced. The propellers are tested separately to test their effect on the flow of a steady stream of water. Flow resistance can now be pinpointed to either hull or propeller faults. destructively with the ship's own wave. Some bulges are being fitted at the stern as well. However, the disturbance from the bulge cancels out the ship's own disturbance only at certain speeds, and it also tends to increase the resistance of the vessel. ee Al Ng Sd \ BME GLE BS 2050 I 7 ——— oS SS SS SSS Se ee," ms — Saas Sa a a Pe hela ee = ~~ —— =

The Testing Tank

The hull is towed along to determine its drag-skin resistance and wave-making. This is carried out in perfectly still water.

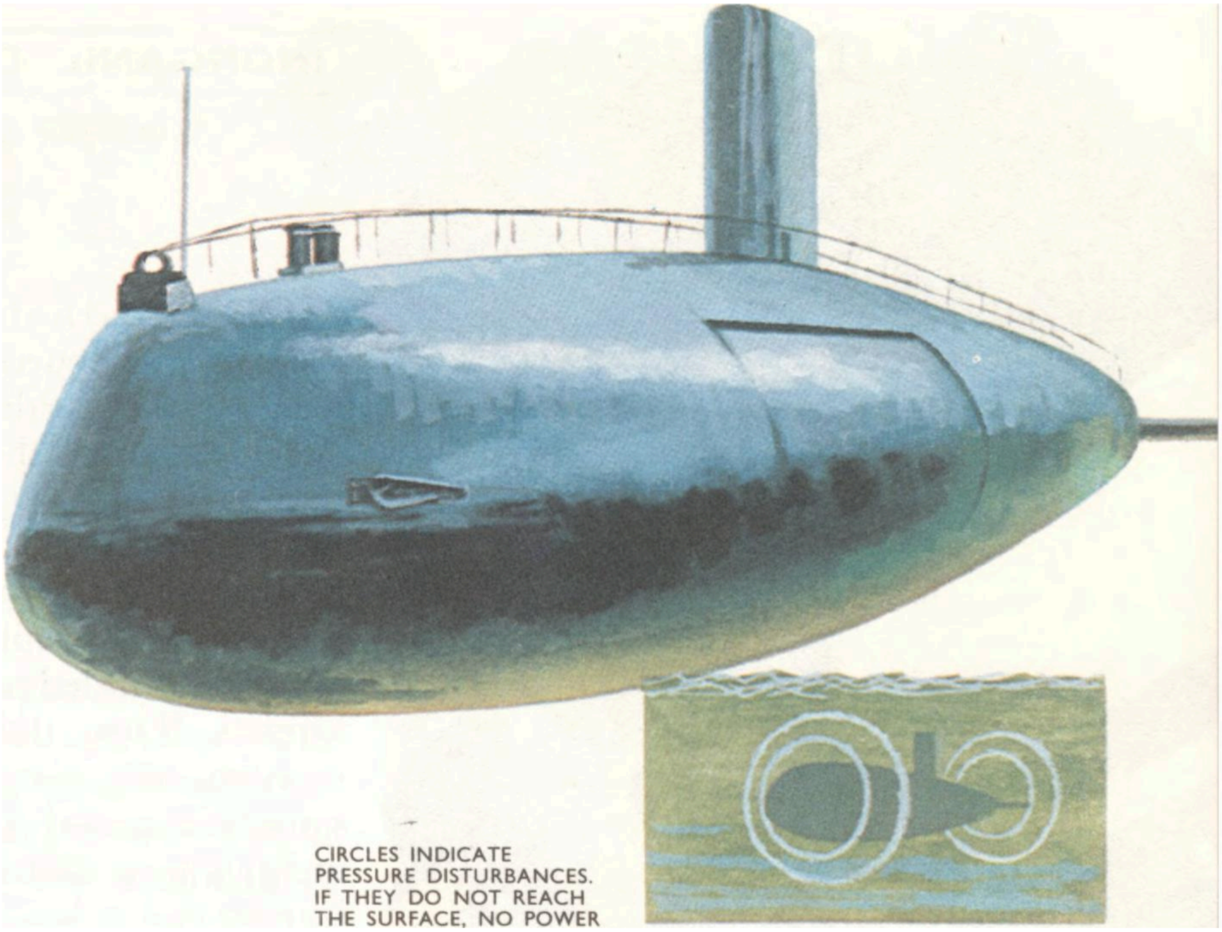
The propulsion system is added and the model moves under its

own power. The overall efficiency of propulsion can now be deduced.

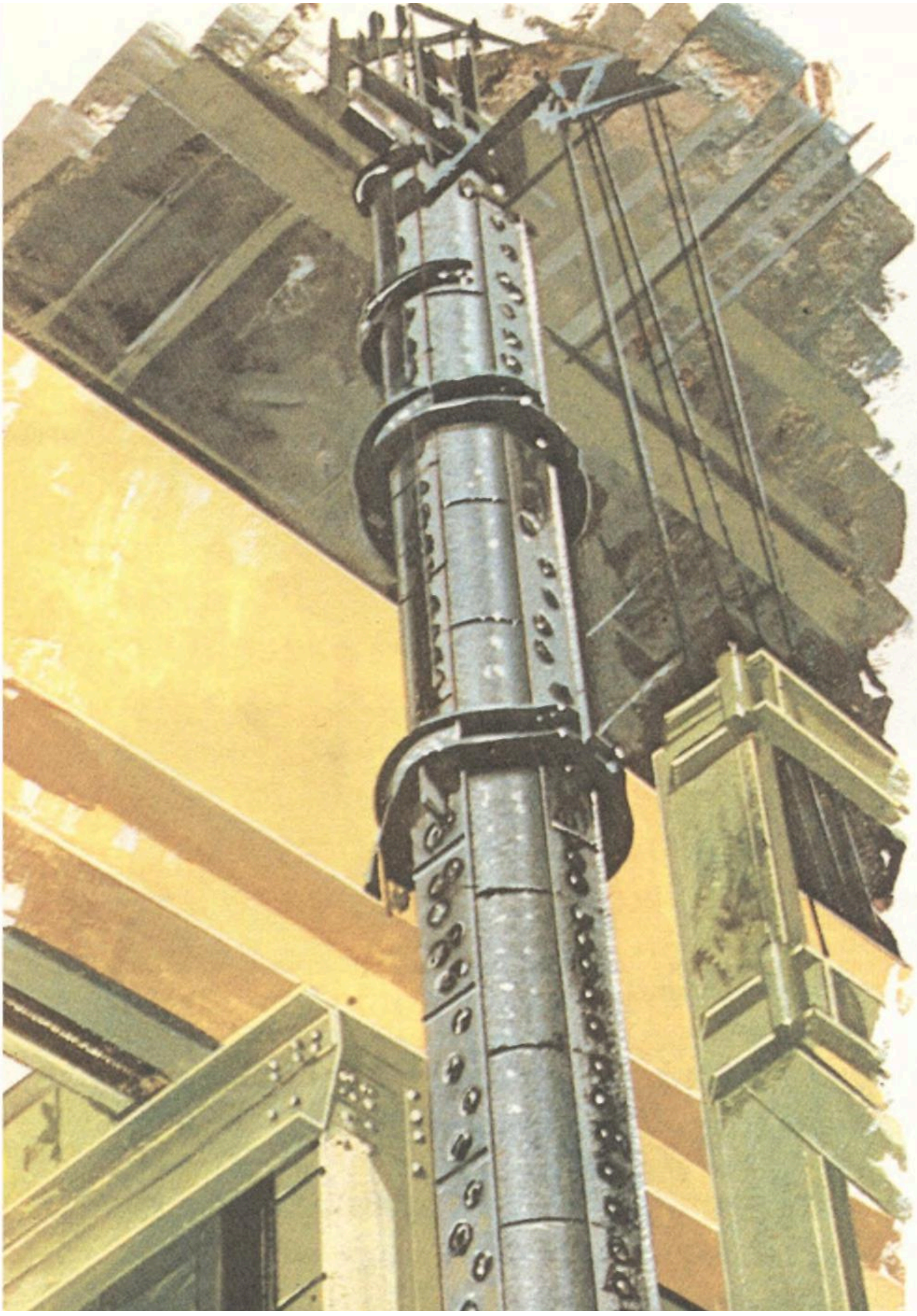


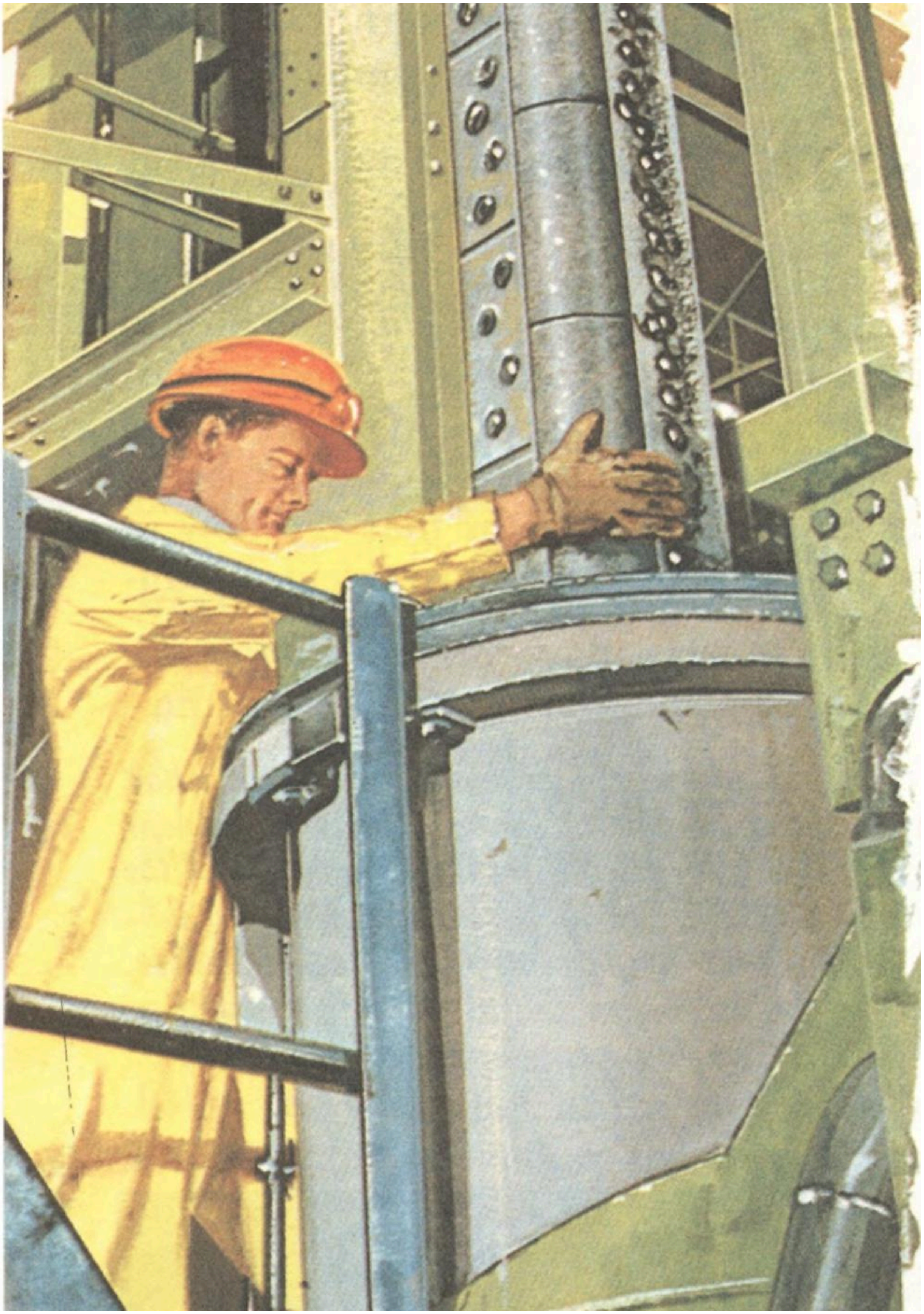
The propellers are tested separately to test their effect on the flow of a steady stream of water. Flow resistance can now be pinpointed to either hull or propeller faults.

Scale Models and Towing Tanks It is unnecessary to build a full-scale ship in order to test some new design—a very good guide to the performance of the full-scale ship can be gained by constructing a model of it, and testing it in a towing tank. Testing speeds in the towing tank are far lower than the corresponding speeds of the full-scale ship. Testing depends mainly on the quantity #. From the point of view of wavemaking, two ships of similar form have exactly the same wave-making pattern when the value of fis the same for both of them. If, for example, a 1,000 foot ship is represented by a 20 foot model, then the wave pattern of the ship at a speed of 20 knots is the same as that of the model at a speed of about 3 knots. The models are usually made of wax so that they can be shaped easily. In the towing tank, they are ballasted so that the water line corresponds to the water line on the full scale ship. Test models are between 10 and 50 times smaller than the actual ship. Model testing takes place in several stages. The model is first tested for drag — the combination of skin resistance and wave-making resistance — by towing it through a tank of perfectly still water. A platform carrying a 'towing carriage' above the water runs on a set of elevated rails along the length of the tank. Instruments connected to the carriage and inside the model give the magnitude of the drag when the model is being towed. In the second stage of the testing the model is powered by its own motor and propellers. By comparing its performance with that in the first test, the overall efficiency of propulsion can be deduced. If the efficiency is much lower than expected, it may indicate a poor design of hull or propeller. In the third stage of testing, the propeller is tested on its own. This takes place in a water tunnel, through which water is circulated at carefully controlled speeds. The model undergoes additional tests for sea-keeping. A machine at one end of the tank produces a continuous stream of water waves. Instruments in the model record the amount of pitch and roll —in other words, how well the model keeps in the water. Parts of the ship which are normally above the water line will dip into the waves during these tests, and the shape of these parts influences the sea-keeping of the ship. One of the advantages of bulbous bows is that they improve the sea-keeping properties of ships. Submarine Streamlining Close to the surface, pressure disturbances at the bow and stern of submarines may reach the surface as waves. But at any appreciable depth this becomes negligible. When streamlining a surface vessel, both wave-making and skin resistance have to be taken into consideration, but only the skin resistance is important in the submarine It can therefore be made to travel at high speeds mor efficiently than any conventional surface craft. High speeds.” S are now possible underwater with atomic-powered engines, and engines which use hydrogen peroxide to supply the oxygen for burning the fuel. Fast submarines are basically pear-shaped. The fore-end CIRCLES INDICATE PRESSURE DISTURBANCES. IF 1 NOT REACH THE SURFACE, NO POWER IS WASTED IN WAVE-MAKING An efficient streamlined shape for underwater operation, the pearshaped atomic-powered submarine. The submarine wastes energy in wave-making only when it ts near the surface. is comparatively blunt, and tapers to the stern. Obviously there has to be one bulbous part to accommodate the engines or the crew, and tests have shown that this is better at the fore-end than at the rear. A bulge at the rear leads to a wake of turbulence, and additional drag. The overall shape of modern submarines is similar to that of the whale. Copying the whale’s shape is hardly surprising since these creatures can move at high speeds underwater. What is more, they and their relatives, the porpoises, move at a greater speed than would be expected from their muscle power, which means that the flow of water over their bodies is approaching the perfect streamlined flow. Towing tests conducted about 200 years ago showed that if there had to be a bulge in a streamlined submarine shape, it was better near the front than the rear. These surprising results demonstrated the importance of turbulence in the wake of a towed object. In submarines, a tapering stern is more important than a tapering bow. But the reverse is usually true in surface craft. NO TURBULENCE CONES ARE BEING TOWED UNDERWATER DIFFICULT TO TOW POINTED END FIRST


CIRCLES INDICATE
PRESSURE DISTURBANCES.
IF THEY DO NOT REACH
THE SURFACE, NO POWER

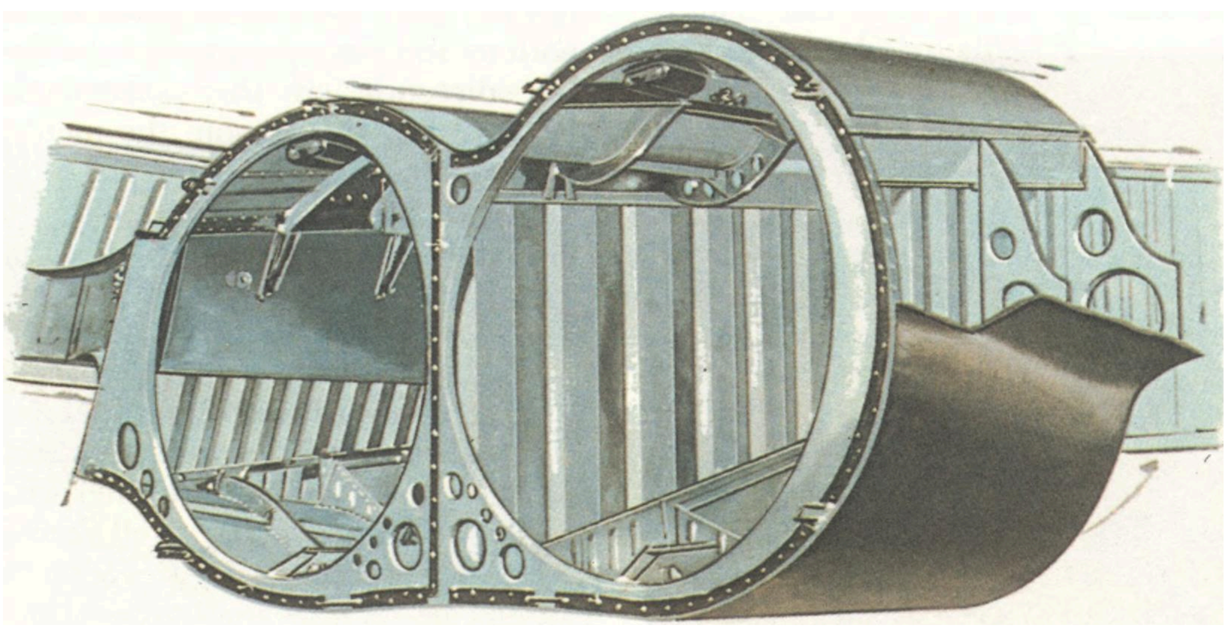
J ww INORGANIC CHEMISTRY ALTHOUGH the metal titanium is the fourth most abundant metal in the Earth's crust, it was the development of the aircraft industry that first thrust it into a position of importance. When first discovered over 150 years ago, it was a problem element which for years puzzled and defeated metallurgists. They did their utmost to economically extract it and make something useful out of it. In fact, the metal was so difficult to extract from its ores that it was not until 1949 that lj, CHLORINE A" ARESEET RUTILE IS TREATED WITH CHLORINE TITANIUM TETRACHLORIDE ~ looked rather like spongy cobbles of coke. The Kroll process is still widely used in America and Japan, but a different chemical method, involving large quantities of sodium, is now used in England, producing dull grey metallic granules instead. In its spongy and granular form the metal is of little use and must be consolidated and have air bubbles and flaws driven out before it can be called useful. This cannot be done by melting and pouring into a cast. Titanium melts at nearly 1700°C, 200° higher than the melting point of steel. At that temperature it would react with the lining of the furnace and absorb gases from the air to make it useless as a TITANIUM MAGNESIUM IMPURE TITANIUM To prepare metallic titanium,rutile is first heated with chlorine in the presence of carbon to an economical method was found. There are two main titanium ores, rutile, an impure form of titanium dioxide, and ilmenite, a mixture of > oxides of titanium and iron. All the ~ titanium metal is obtained from rutile while titanium compounds are manufactured from ilmenite. The method worked out for titanium metal in 1949 by the American W. J. Kroll involved converting the titanium in the ore into titanium tetrachloride TiCl,. This is then reduced to the metal by reacting this compound with . magnesium. The metal produced ! After being compressed to form a huge electrode, the titanium is loaded into a ~ crucible. The crucible has its air withdrawn and the electrode slowly melts and con© solidates. yy form titanium tetrachloride. This is then reduced to the metal by heating with magnesium. structural metal. Granules of raw metallic titanium, sometimes mixed with other powdered metals to make alloys, are thoroughly blended and fed into a 2500-ton press which squares the powder into a briquette. The briquettes are then welded together to make an electrode 12 feet long and weighing almost a ton. The electrode is suspended from the top of the furnace while a water-cooled crucible is clamped on at the bottom. Air is drawn out, an arc is struck between the electrode and a small quantity of titanium powder in the crucible base. The electrode slowly melts to form an ingot. The melting is repeated with the whole operation handled by remote control. High frequency sound waves (ultra





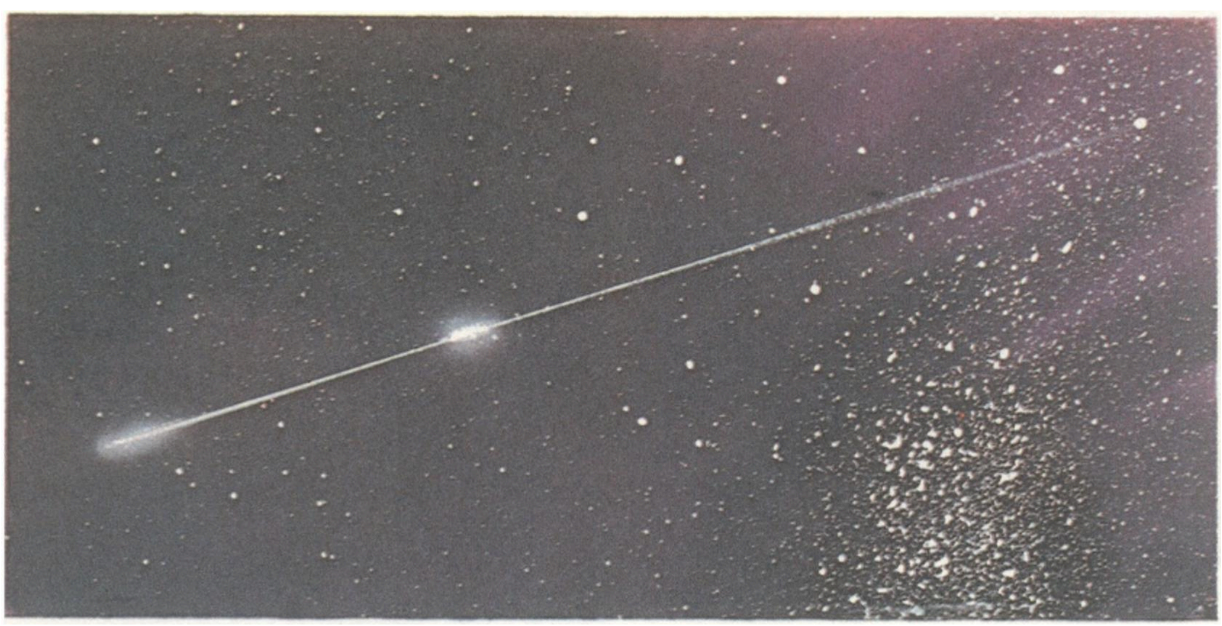

2052

sonic waves) are used to test for flaws. This is an echo sounding technique. Flaws inside the metal act as mirrors for the beam, preventing it from passing through the metal. No signal on the far side indicates a flaw. Aircraft Engineering The aircraft industry requires lightweight alloys which can withstand the stresses imposed by ultra-high-speed flight and titanium provides the answer. Its density is only 60% of the density of steel while it keeps its strength at temperatures far higher than those considered safe for aluminium alloys, the other lightweight alternative. The aircraft industry uses titanium for its turbine compressor blades, exhaust shrouds and hot air ducts, also on the leading edges of the wings where it can resist erosion from the air rushing past. Because of its high resistance to corrosion by acids etc., the metal is also being used for manufacturing corrosion-free vessels and pipes for the chemical industry. On a much smaller scale, but for the same reason, surgical implements such as internal splints and screws used for fixing bones together, clips and surgical nails made of titanium are gradually taking over from stainless steel. Radiation Shielding Nuclear power stations use titanium for many of their internal components Titanium is used for surgical repairs because it does not corrode inside the body. because of the ability of the metal and its alloys to shield radiation. The irradiated metal quickly loses any radioactivity,making the parts easy to handle and easing the maintenance of the reactor. White Pigment Many white paints and inks owe their whiteness to the pigment, titanium dioxide, TiOg, the only compound of titanium of any real importance. White plastic floor coverings and white rubber are white in colour because this compound is incorporated in them. Textiles can be sprayed with it to counteract unwanted lustre and enamels and tile glazes can be treated with it to regulate the colour, opacity and gloss. There are two different ways in which titanium oxide is used by the paper industry. It can be incorporated © during the making and its particles held throughout the body of the sheet to reflect the light and make the paper appear whiter, or it can be painted on Titanium is used by the aircraft industry because of tts strength, resistance to corrosion and light weight. Pure titanium is used for the firewalls on the engine bays of this comet. as a surface coating. It is usual to coat the thicker papers with the oxide, but for airmail paper where it must be flimsy but not transparent, the oxide pigment is spread through the paper pulp during the making. White waxed wrapping paper is only white because of the titanium oxide added to it. The extraction of the metal titanium and the manufacture of its See | Titanium as a a Transition Metal s62 7s as As we. sieceendicare det' to right in the cereale table, the Ss gain an extra electron, one at a time their outer shells to buile of eight electrons. But which might reach 18 = Titanium is one ofa group of nts— the transition group. " These all have two. electrons in the outer shell but the inner shell has penweet he electrons and 18 electrons. oxide are quite independent processes. The pigment is not made from the metal, but also has its starting point in the titanium ore ilmenite. ; The pigment is made from ilmenite by crushing the ore and dissolving out the titanium with concentrated sulphuric acid. When the solution has been boiled down and cooled, the iron that also dissolved crystallizes out and can be separated. Concentrating the liquid even further brings out the titanium as hydrated crystals of titanium sulphate. These crystals are filtered out and washed before being fed into a rotary furnace where sulphurous gases are driven off, leaving impure particles of titanium dioxide behind. After purification and grinding to the right size they are ready to be mixed with paper pulp or paint. 2053



ASTRONOMY METEORS and METEORITES OVING at 45 miles per second, a fragment of stone, no more than a millimetre in diameter, strikes the Earth’s atmosphere. The friction of the stone against the atmospheric gases is intense. The tiny stone’s surface begins to glow with heat. The result? —a shooting star, a bright line of fire which flashes momentarily across the skies. Such fragments of stone are called Fossil Crate rs The use of aerial photographs in geological surveying has led to the detection of a number of perfectly circular structures on the Earth’s surface. Though partially obscured by sediments, the structures on inspection have been shown to be old craters probably formed from ancient meteor_ ites. A dozen or so have been found in ’ Canada for instance — one measuring 20 miles across. Dating of the sediments give the age of some of the craters as at least 500 million years. Perhaps the heavy meteoritic bombardment believed to account for the Moon’s pitted surface also affected the Earth at some stage in its history. The craters on the airless, waterless Moon would be preserved while the majority of Earth’s structures would be erased by weathering processes. meteors. Like the Earth, and other planets, meteors revolve in orbits about the Sun. A diameter of one millimetre is about average size. Some, a centimetre or so in diameter, produce exceptionally brilliant shooting stars. Very occasionally a ‘giant’ meteor — between 1 and 10 centimetres in diameter — strikes the atmosphere giving the spectacular ‘fireballs’ and a few — the so-called bolides — mark their entry by loud explosions. On almost any night several meteors can be seen as shooting stars each hour. Over the entire Earth, something ‘like 25 millions of them have been estimated to fall every 24 hours. And these are only the bright, observable ones; taking into account the vast numbers of minute fragments which go unnoticed, the total swells to the staggering figure of 400 million. Despite this colossal number, the actual increase in the Earth’s mass is only about a ton a day — negligible compared with the Earth’s total mass. Meteors are, of course, far too small to reflect sunlight like the Moon and the planets. Consequently, they can only be detected when they enter the Meteors are only visible when they enter the Earth’s atmosphere. The friction causes them to glow with heat. Usually they appear faint at first, become brighter, then fade finally ending in a flash. From the direction and velocity of meteors their orbits can be estimated. Sizes can be calculated from the brightness of the flashes. 2054 The Arizona Meteorite Crater (the

Barringer Crater) is about three-quarters of a mile across- and 600 feet deep —the largest meteorite crater on the Earth. The age of the Earth's atmosphere and become incandescent. The height at which they begin to glow is about 65 miles above the Earth's surface. Usually they are completely burnt out 50 miles up, and the trail of light ends abruptly. Very bright fireballs, however, may be visible down to distances of only fifteen or ten miles above the Earth. At certain times of the year, particularly large numbers of shooting stars can be observed — the spectacular meteorite streams. A 'stream' will occur when the Earth in its orbit about the Sun, passes through the orbit of an accumulation of meteors. For instance, between July 27th and August 17th of each year, the Earth is bombarded with meteors coming apparently from the direction of the constellation Perseus. For this reason, the meteors making the shower are called the Perseids. The streaming effect — the apparent divergence of the shooting stars from a single point (or radiant) is one of perspective. Each individual meteor is travelling in an orbit parallel to its fellows. Just as the parallel lines of a railway track seem to issue from a distant point, so do the parallel orbits of the meteors. Other well known meteoric storms are the Orionides (seen coming from the constellation of Orion between October 2oth-2!st)



In 1961 two stony meteorites — one from southern France, one from Central Africa—showed what appeared to be the fossilized remains of very simple organisms. The ‘organisms’ rather resemble our own algae (simple plants). Some are circular with double walls, some shield-shaped, some even cylindrical with heavy sculpturing at. the surface. All are microscopic in size. Meteorites are the only extraterrestrial objects that can be handled. Do these structures mean that life exists (or existed) on some other planet? Whether the finds are in fact organic in origin is difficult to say. Maeee. 2 ne a ta Some slight confirmation comes from a traces of hydrocarbons also detected NEEDED in the meteorites. Such hydrocarbons Across THE could have been synthesized by algallike organisms. — : CRATER a ied crater is unknown. Estimations made from meteorite, which probably exploded on hitthe rate of eroston and the rate of rusting of ting the ground, must have weighed more fragments of iron meteorite suggest some- than a million tons. Another large crater is SRE sun thing in the order of 50,000 years. The the Canadian Chubb Crater. : ert , and the Geminids (issuing from the orbits at exactly the same time. For OF EARTH OO xt ; Pie constellation of Gemini between De- example, the Draconids only appear : ¥ * cember 12th-13th). every 13 years. In the examples given above, the What are the origins of meteors? meteors are widely dispersed through- Where did they come from ? Certainly ARS: out their orbit. This means that the most of them seem to be closely siecal meteor stream will occur whenever associated with those small wanderers EVENING the Earth’s orbit and the meteors’ of the solar system, the comets. For ‘*" orbit coincides. But sometimes the example, the Draconids have exactly meteors are condensed into a restricted the same orbit as the Giacobini-Zinner _ FAST iced akey cloud. Then the streaming effect will | comet and appear to be debris thrown — ON” NIGHT ss saice telnet occur only at intervals of a few years, off by it; the Leonids revolve in the se : whenever the meteorites and the wake of the Tempel comet, first disEarth reach the intersection of their covered in 1866; and the famous In tropical regions after sunset and before Haley’s comet is associated with ite sunrise, a glow may be seen in the sky, A ™eteor showers. An actual on bright cone on the horizon, the glow fades recorded of a known comet (Biela’s away towards the zenith. This zodiacal] comet) which first split up into two light zs believed to be caused by meteoritic | and then broke up altogether, leaving $ $ dust circling in an orbit close to the Sun. acloud of meteors. : : Meteorites > METEOR It is fortunate for the inhabitants of 4 8 OreIriN the Earth that meteors do, in the main, a FP gen 3 burn out in the upper reaches of the $ e i atmosphere. Otherwise the Earth’s surface would be continually bombarded with very rapidly moving particles. Occasionally, however, larger bodies actually reach the surface still intact; not all the solid material has vapourized. These are the meteorites. They range in size from the dimensions of a pebble to great blocks 2055

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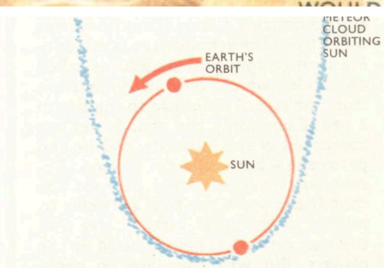
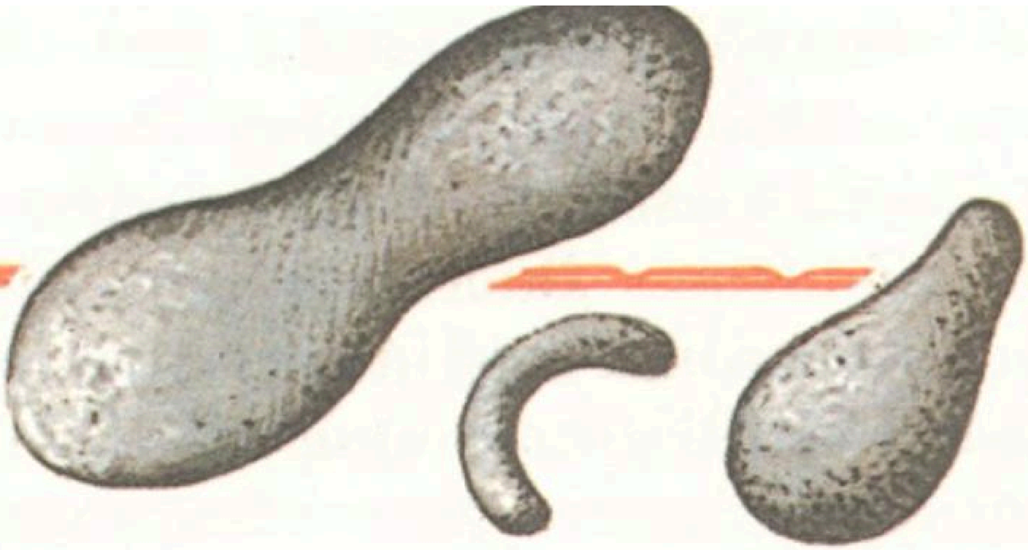


Meteor streams occur when the orbit of the Earth intersects the orbit of a meteor cloud.

CRUST OF CORE OF IRON AND NICKEL ‘CENT “ ND DID NOT SPIN The two main classes of meteorites the irons (below) and the stones (above) can be related to zones of rock making the Earth. This is strong evidence that the meteorite originated from the break up of a planet that structurally resembled the Earth. TS A. "Whee erates ca three-quarters of a mile in diameter and 600 feet deep. Despite many attempts to excavate the meteorite, no single large mass has been recovered. Perhaps the meteorite exploded on striking the ground; certainly the whole neighbourhood was found to be sprinkled with meteoritic fragments. More recently, two meteoritic falls have occurred in Siberia, one in 1908 and the other in 1947. The 1908 meteorite devastated wide areas of forest. Usually meteorites fall into one of two classes—the stones (or aerolites), and the irons (or stderites). The stones make up about 90% of all known meteorites. They consist largely of silicates and are not very different from the rocks making the crust of the Earth. The irons are metallic bodies, consisting of mostly the element iron (usually about 90%), some nickel (usually about 8°) and traces of other elements. A few meteorites — the siderolites — are composed of both silicates and iron. Origin of Meteorites The majority of meteorites appear to have a different origin from meteors. They are as a rule larger, and do not have close associations with comets. Further, perhaps because of a higher overall density, they are not so fragile as meteors and not so readily volatilized on entering the Earth’s atmosphere. No matter what the type of meteorite, all show signs of a previous molten state. Many of the iron meteorites in fact have a peculiar crystalline structure not known in any terrestrial rocks ; from laboratory experiments, the structure seems to indicate cooling from a molten condition under very intense pressures. Where would this \MOOTH \URFACE several tons in weight. Though meteorites were reported as long ago as | 644 B.c., their true celestial nature | was not finally accepted until 1803. The largest known meteorite fell in South Africa during prehistoric times. It weighs 60 tons and must have originally weighed about 80 tons. Another, of 33 tons, has been recovered from Greenland. Probably the most massive body of all reaching the Earth’s surface was the meteorite responsible for the — Crater, | @ @0 @ | _ 9 °D Nog 92 gd CHONDRULES —--NONE ARE LARGER Chondrites and Achondrites THAN THE SIZE OF A PEA The stony meteorites or aerolites are further subdivided into chondritic and achondritic classes. The chondrites differ from the achondrites in containing strange spherical structures called chondrules which vary in dimensions from the size of a pinhead to the size of a pea. Commonly the chondrules are composed of the minerals, olivine and pyroxene — both silicates of iron. Recent work suggests that perhaps the chondrules were the first particles to solidify in the solar system, and that their accretion has helped build up the planets. But why then are no chondrules known to exist in terrestrial rocks? If, in fact, they ever were present, the processes of erosion would long ago have destroyed them. 2056 pressure come from? Meteorites are as a rule fairly small, not capable of providing such pressures themselves. The suggestion follows that meteorites were once part of a much larger structure —a planet perhaps, within our own solar system. The composition of the meteorites provides additional evidence. A strong similarity exists between the various types of meteorites and the various zones of the Earth's crust. The stony meteorites for instance could coincide with a rocky outside layer, the irons with a metallic (high pressure) core. Looking for a site for a possible original planet is not a lengthy job. Between the planets Mars and Jupiter, revolving in an orbit about the Sun, are numerous fragments of rocky material — the asteroids. The asteroids almost certainly belonged to a former planet that broke up some time in the remote past. Perhaps the majority of meteorites have the same origin. Certainly the radioactive dating of meteorites — giving an age of about 4000 million years — is approximately the same as the Earth and probably the other planets of the solar system. Tektites Tektites are dark, glassy lustrous stones found in many parts of the world. They range in size from less than an inch in diameter to more than five inches. Shapes are variable — discs, orbs, rods, even dumb-bell shaped stones are known. Their glassy nature suggests rapid cooling from temperature as high as 1700°C. Tektites are usually found in groups centered around certain areas. The age of the groups can often be discovered by dating the rocks in which the tektites occur. Some groups are 45 million years old. Tektites were formerly thought to be a type of meteorite. But today i" are believed to be in fact terrestrial. Their origin seems a dramatic one; they oo to be the solidified droplets of molten rock splashed up into the air when giant meteorites have torn into the Earth.



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BIOLOGY SEXUAL REPRODUCTION IN SPIROGYRA -A SIMPLE PLANT. GAMETES ARE THE SAME SIZE CHROMOSOMES OF (DIPLOID STATE) aa & ZYGOTE - CHROMOSOMES” OMBINE TO GIVE A COMPL The great advantage of sexual reproduction is that there 1s continual exchange of genetical information. The number of chronosomes in the gametes is halved, though each single chronosome contains elements from the original pair. When gametes combine, the zygote produced again has a complete set of chronosomes. Usually in animals gametes divide thewr labour — large eggs carry food, sperm remain small but are mobile. Reproductive Methods REPRODUCTION i is the most important function of any living species; failure to reproduce means that the species will become extinct. In response to this biological necessity a number of methods of reproduction have developed. Amongst higher plants and animals sexual reproduction is the most widespread and successful method. At some stage, specialized reproductive cells (or gametes) develop. One gamete must combine with another to form what is called a zygote. The process of combination is called fertilization. Gametes must not only be capable of coming together. Between them, they must provide enough food for the new individual to develop. Usually there is a division of labour; one type of gamete — the female cell or egg — remains stationary but contains a large food supply. The other type — the male or spermatozoan — carries a minimum supply of food but is mobile and moves towards the egg. Commonly, in animal species, separate individuals bear opposite sex cells. Male and female animals are recognizable and the species is said to be sexually differentiated. When the two types of gamete are produced by the same individual, the animal is an hermaphrodite. Earthworms and flatworms, for example, are hermaphrodites. Usually elaborate structures are present preventing selffertilization. In plants it is far more common for a single individual to produce both types of gamete. The gamete differ from the other, body cells (the somatic cells) in the number of chromosomes it has present in its nucleus (the chromosomes are the rod-like structures containing the genes or factors controlling inheritance). The non-sexual, body cells each have a paired set of chromosomes, the number of pairs varying according to the animal. Man, for instance, has 46 chromosomes in the cells of the body (or 23 pairs). But the specialized sex cells, by a process called meiosis, come to have only a half a set. For instance, Man's sex cells have only 23 chromosomes However, during the process of fertilization, the male and female sex cells combine forming a single zygote. The zygote, receiving two half sets of chromosomes, comes to have a complete set. By division it gives rise to all other cells of the body and they likewise all have full sets of chromosomes. There are a few creatures in the animal kingdom which may reproduce by a process called parthenogenesis. The female produces eggs in the normal way but the nucleus of the egg retains its full set of chromosomes. Consequently no fertilization is needed. Examples of this strange method of reproduction are found in the aphid (or plant-louse), daphnia (the minute crustacean the water flea) and rotifers. fh) USUALLY THERE Is A DIVISI oe Gerace IN GAMETES BUT MOBILE y, IN GAMETES, NUMBER OF CHROMOSOMES IS HALVED (HAPLOID STATE). EACH SINGLE CHROMOSOME CONTAINS PIECES FROM THE ORIGINAL PAIR wai) In plants, a similar method may occur. Though outwardly a sexual mechanism may be indicated (that is, flowers develop and pollination occurs), the female sex cells may never be fertilized. Commonly they retain their full set of chromosomes and the egg develops directly. This process in plants is called apogamy. A further type of reproduction is the asexual method which occurs mainly in plants but also in a few simple animals. No specialized sex cells appear. New individuals develop directly from the Bascal by cell evasion Neadoisth 2057 EGG SWOLLEN WITH FOO o- | SPERM ARE SMALL ei j 4 \spenn ¢

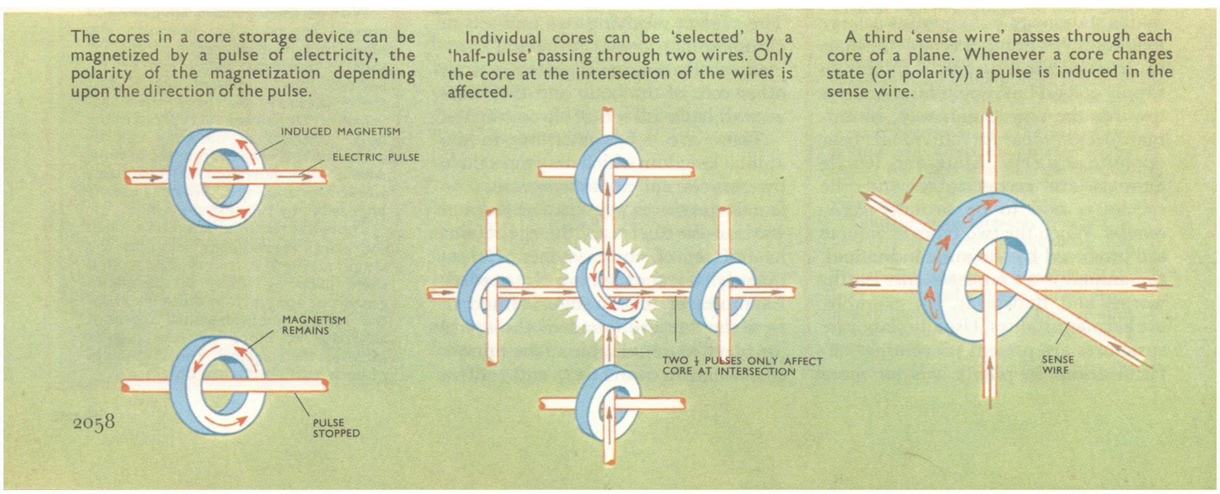
Advantage and Disadvantage

The great advantage of sexual reproduction is that recombinations of genes are continually occurring. At each fertilization different genetical material comes together. The variety of animals likely to be produced is thereby increased for it is the genes that largely control development. It follows that the chances of an animal appearing with advantageous adaptions is also increased.

With parthenogenesis, apogamy and asexual reproduction, however, the offspring will resemble the parent, for exactly the same genes will be inherited. No variations occur. But these methods do have their advantages. Unlike sexual reproduction, large numbers of offspring can be produced in a very short space of time. Also a *single* individual may colonize a whole new area, a feat which is usually impossible by the sexual method.

Though it is common to find animals and plants reproducing solely by the sexual method, it is very rare to find them breeding solely by partheno-genesis or apogamy. In these instances the sexual and non-sexual methods alternate at more or less regular intervals (*heterogamy*). The advantages of both mechanisms are obtained.

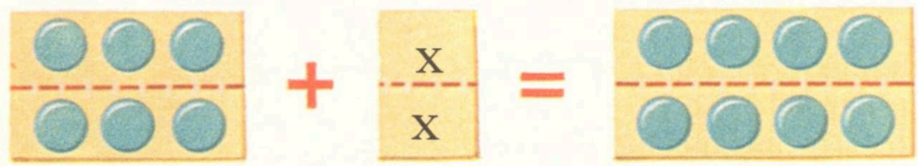
ELECTRONICS HE essential difference between a calculating machine and a computer, is the possession of memory devices enabling information to be stored. The operator of a calculating machine has to remember, usually by jotting down on paper, any intermediate result which he will require later. Sometimes the machine will print the results itself, but it cannot read its own writing, and so cannot Information Information is most conveniently stored by any mechanical, electrical or electronic device: in the form of binary numbers. In ordinary arithmetic we use ten different symbols to represent numbers, using one of them to represent a number below ten, two for numbers below one hundred and so on. Being accustomed to numbering to the base ten it is difficult to imagine using any other base, though in fact - any number could be used. Computers use the binary system, with the base two, and twosymbols | and0. Numbers below two can be represented by just one binary digit, below four by two of them, below eight by three, below sixteen by four and so on. Thus binary 1010 (read it ‘one-O-one’) is equal to decimal ten. use this as a memory. The operator must also remember what to do with the information, or data, to be worked on, which numbers are to be added, which subtracted and so on. A computer can remember not only intermediate results, and refer back to them later, but its instructions, or programmes. Some computers are even equipped with such large memories that they can remember complete sets of accounts, or of totals of stocks held, and can keep these records constantly revised as transactions take place or stock is used or replenished. The devices which can be used as memories are many and various, but three kinds are in very common use in modern computers. These are core storage, magnetic drum storage and magnetic disc storage. Magnetic tapes, similar to those used in ordinary tape recorders, can also be used for storage, the computer being able to run the tape forwards or backwards to find the information required. Tapes are also used to feed in data for processing, or for the computer to write out its results. Core Storage These devices are made of thousands of separate very simple devices, each capable of storing a signal pulse. This represents a one or a zero, and is called a binary digit (or bit). Each device consists of a tiny ring — or core — of special material such as ferrite, which has a strong magnetic property. These cores are strung on wires to form flat carpet-like planes, and several such planes are stacked one on top of the other. Each core can be made into a magnet by passing an a X ak ak ak SINAN NGAS INF SENSING SHINAI NINE MN GMI NENA SINAN ENEXS a at de vd SEMEN NAS SNE NENGINS The cores of a random access storage plane are strung out on wires through which pulses can be passed to ‘read’ the memory or to ‘write’ information into it. Each tiny core can store one binary digit. Several planes are stacked on top of one another to form a memory unit of the required size. electric current through one of the wires through the ring. The direction of this magnetization (or its polarity) depends on the direction of the current, and the core remains magnetized, even when the current has stopped. If the current is then reversed, the magnetic polarity of the core is also reversed. Only a brief pulse of current is necessary to magnetize, or reverse the magnetism of a core, but it must be fairly powerful. In the plane of cores, each core has two main wires passing through it. These wires are arranged in ‘columns and rows’ like a table of



logarithms, with a core at each intersection. A “half pulse’ can be sent through one column and one row, which on its own is insufficiently strong to affect a core, but at the intersection the combined power of the two half pulses is sufficient. So in a plane of, say, 100 cores only 20 wires are required to be able to ‘write’ into the memory up to 100 bits, by using the wires in pairs. In this form, the memory is rather useless, because there is no means of reading out the stored number. However, a third wire, called the sense wire, passes through every core of the plane. Whenever a core reverses its polarity, an electric pulse is induced in the sense wire by the changing magnetic field. To read out any particular core, a zero pulse is sent through the pair of main wires which cross at it. If that core already contained a ‘zero’, no pulse would be received on the sense wire. If it contained a ‘one’, then there would be a pulse on the sense wire. Reading out from core storage would normally make the memory ‘forget’ everything it contained. This is normally remedied by automatically ‘rewriting’ the information into the store. It does this by sending a ‘one’ pulse to each core whose contents have been read, while a fourth inhibit wire, which passes through every core of the plane cancels out one ‘half-pulse’ for each core from which no pulse was received in the sense wire (i.e. which contained a ‘zero’). Drum storage uses a drum (made of non-magnetic material, e.g. copper) coated with a magnetic material with similar properties to that used to coat magnetic tapes. The drum is rotated ‘at a steady, and usually high, speed. A read/write head, which is really a tiny electromagnet, with a very small gap between the two poles, in which the magnetic field is concentrated, is placed very close to the moving surface of the drum. A pulse of electricity in the write coil of the head will magnetize a tiny patch of the surface coating of the drum, and the polarity of the magnetism depends on the direction of the pulse current in the ‘write’ coil. Reading is simply a reversal of the process, the passing magnetic patch inducing a pulse in the read coil, the direction depending on the polarity of the magnetic patch. The drum is divided up into narrow columns, each one forming a band around the drum, and either a separate read/write head is provided for each column or set of columns, or the head moves along to select the required column. There may be hundreds of columns on one drum and each column can store several thousand characters, so the amount of information stored can be very large, in the region of a million characters on a medium-sized drum. Disc storage devices are very similar in principle to drum storage, except that the columns are on the faces of discs, similar to gramophone records, and that the grooved track of a record is spiral, not a series of concentric tracks. Many discs can be stacked one on top of the other, and both faces used for storage. The read/write heads are arranged on a long arm which passes between the discs to scan the faces. Very large capacity memory is obtainable with disc storage devices; it is this type which is used to 'store' complete company accounts or stock totals for instant reference. The access time of any memory device is of great importance. This is the time taken to actually write or read a digit into or out of the memory. The access time for magnetic tape can be several seconds, while for core storage devices, the access time may be a few micro-seconds (i.e. millionths). It is largely because computers are able to work at such high speeds that they are so useful, for they can do thousands a ss = a 7" Disc storage devices are capable of remembering vast amounts of information and of keeping it continuously up to date. The surface of each disc is diwided into concentric tracks, any one of which can be reached by the finger-like read/write heads. of simple calculations in a single second. New memory devices with even shorter access times are being developed. Among them are those using thin film techniques in which extremely thin films of magnetic materials, only a few molecules thick, are deposited on to various substances. Several films can be laid together to form cells which can, like ferrite cores, be switched to one or other of two states. Their extremely small size enables them to change magnetic state in very short times and with small power requirements. Information is stored on surface of a drum. storage device, in the form of tiny magnetic patches formed by a 'write' current passing through the read| write head. The polarity of the magnetization depends upon. the direction of the current. As these tiny magnets pass the head, they induce pulses in the 'read' coil of the head. 2059

MATHEMATICS PRIME NUMBE _ = * e ‘ “) LRS J CANCELLING "THERE are several equivalent ways of writing some numbers. For example, 4 can be written as 2 x 2. 6 can be written as 2 x 3. 2 and 3 are ‘called the factors of 6. But 2 and 3 themselves cannot be split up into factors (except themselves and one, 3=3-x% 1).. They are said to “be prime numbers. The number 6 can be represented Rectangular numbers 3 rows 2 columns 2 rows 3 columns 2x3=3 x2. 2 and 3 are both factors of 6. It does not matter which factor comes first. This is known as the commutative property of numbers. by 6 counters. 6 counters arranged haphazardly can stand for 6, but if they are arranged in 2 rows of 3 or 3 rows of 2 their factors are immediately obvious. Any number which can be arranged in a rectangle like this is called a rectangular number. 4 can be arranged either all in one row, or as 2 rows of 2. The result is a square. Square numbers are special cases of rectangular numbers. It is impossible to arrange prime numbers in a rectangle. There are always either counters over or counters missing. The prime number has only two factors, itself and one. Arranging 6 counters in a 2 x 3 rectangle demonstrates a very important property of most kinds of numbers. 2 x 3 is exactly the same as 3 x 2. It does not matter which of the two factors we take first. Sometimes a number appears in an equation as one number divided by another. For example 6 This number is obviously equal to 2. The point is that when one or both of the numbers can be split up into factors, they may be found to have a factor in common (a common factor). 6 3 x 2 3 3 3 is common to both the top and the bottom. When this happens, it is possible to divide both top and bottom of the fraction by the same factor (cancel out the common factor, leaving only 2). This is done automatically for a very simple problem like 6-over-3. Cancelling is useful when the numbers involved are cumbersome. It is a way of simplifying an equation. Dividing both sides of an equation by the same number The equation reads: 6 + 2x = 8. It is not in its simplest form, because 2 is a common factor throughout. So both sides of the equation can be divided by 2. + 000+ x|= 0000 The equation: 3 + x = 4is equivalent to the above equation. The double-headed arrow is the symbol for 6 is equivalent to’. 2060 Is | prime? It also happens to be a square number. | is rather a special case, but it is usual to regard it as prime. 2 is prime. It has no factors other than itself and | —it is impossible to arrange 2 counters symmetrically in anything other thana2 X | array. 3x il 3 is also prime. 4 is not prime. It is a square number. There are two alternative symmetrical arrangements —4 X | or 2 x 2. 5 is prime. There is always an odd space left over when 5 counters are arranged in anything butastraight line. 7 is prime. 9isasquare number. 9x lor3 x3

Dividing both sides of an equation by the same number



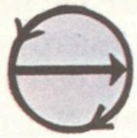
The equation reads: $6 + 2x = 8$. It is not in its simplest form, because 2 is a common factor throughout. So both sides of the equation can be divided by 2.

The equation: $3 + x = 4$ is equivalent to the above equation. The double-headed arrow is the symbol for 'is equivalent to'.

MATHEMATICS The Irrationals RATIONAL numbers include fractions and decimals as well as whole numbers. Rationals are numbers which can be defined as a ratio, or as one whole number divided by another whole number. For example, sis a rational number. So is 1-337. An alternative way of writing this second number is: 1,000 Both 1,337 and 1,000 are whole numbers (they belong to both the set of natural numbers, and the set of integers), sO 1-337 fits in with the definition of rational numbers. In the ordered pairs notation, 1-337 would appear as: (1,337, 1,000). One reason why rational numbers are so defined is to distinguish them from another important set of numbers, the zrrationals. It is hardly surprising that these are the set of numbers which cannot be defined as ratios. They appear only as mathematical expressions with an infinitely large number of terms. mt (pie) is one very important irrational number. The values of 22 and Rational Numbers Some of the rational numbers between 0 and | are mapped on to the number line above. Rational numbers can be written either as fractions, decimals or ordered pairs. Irrational Numbers 3-142 used in simple calculations involving m are only approximations. mis the ratio of the perimeter of a circle to its diameter. Rational numbers could be found by comparing markings on number lines—for example a number line marked in feet with a number line marked in yards to obtain the set of thirds, (1, 3)> (25 3)s (3s 3)s (453) ++ This method does not work for irrationals, because the position of any irrational on a number line can never Some important irrational numbers be pin-pointed accurately. All that can be said of an irrational number is that it is bigger than a rational number to its left on the number line, and smaller than a rational number to its right. m lies between 3-14159 and 3-14160. N is definitely there, between these two markings, but it is a fruitless task narrowing the range much further. No amount of rangenarrowing will ever give an exact value for N. m is associated with the circle, and thus with any kind of cyclic, to-and-fro motion e is approximately 2-178. It is the exponential, and the basis of natural, or Naperian logarithms. 1-414 is only an approximate value for the side of a square of area 2 square units. The cube root of 2 is also irrational. Its approximate value is | -260. 2061

Some important irrational numbers

π

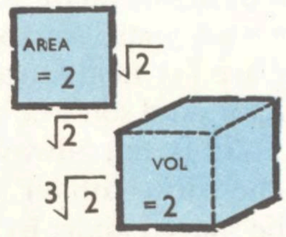


π is associated with the circle, and thus with any kind of cyclic, to-and-fro motion

e

e is approximately 2·178. It is the *exponential*, and the basis of natural, or *Naperian* logarithms.

$\sqrt{2}$



1·414 is only an approximate value for the side of a square of area 2 square units.

$\sqrt[3]{2}$

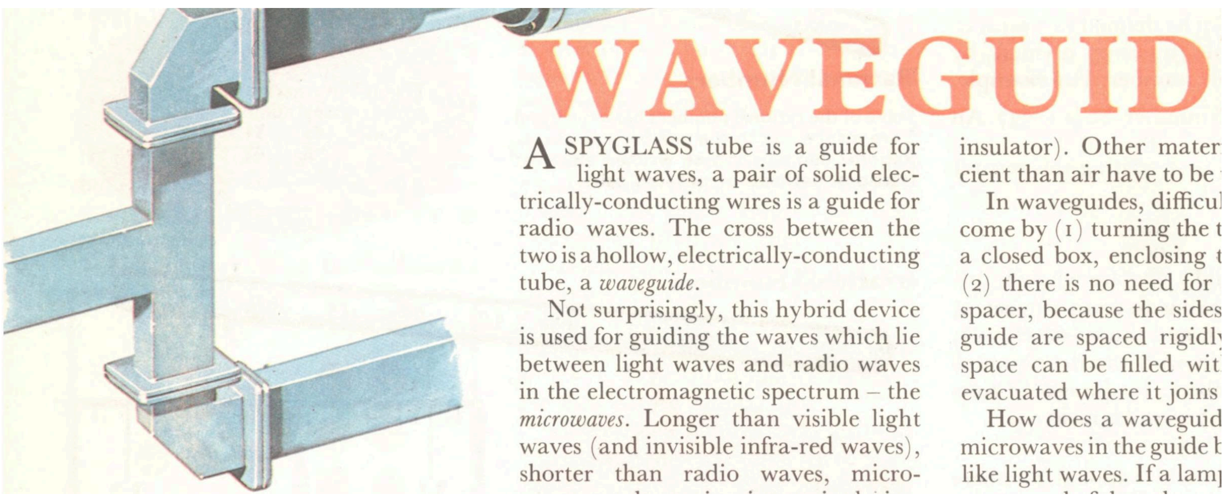
The cube root of 2 is also irrational. Its approximate value is 1·260.

Filling in the spaces on the number line

Each rational number can correspond to a mark on the number line. If all the rationals between 0 and 1 were marked by a line, the number line would be completely covered in marks, because the number of rationals possible is infinitely large.

But the number line appears covered only because of our clumsiness in drawing the marks – theoretically they should have no width at all. Empty spaces remain between the rationals, no matter how many of them are drawn in. The spaces are filled in by irrational numbers. They are in theory essential if the complete line is to be used for measuring lengths, in terms of any of the common units – the yard, foot, metre, or centimetre.

ELECTRONICS This waveguide ts part of the aerial feed of a microwave telephone and television link. Microwaves can travel around properlyshaped bends. 1 Below : the bent wave-guide extracts some of the energy from the straight guide. The guides are coupled by transverse slots. measuring aertal direction patterns. Below : the waveguide horn acts as an aerial Sor incoming or outgoing microwaves. It can be used to feed a larger aerial, or fo PHASE OF A SPYGLASS tube is a guide for light waves, a pair of solid electrically-conducting wires is a guide for radio waves. The cross between the two isa hollow, electrically-conducting tube, a waveguide. Not surprisingly, this hybrid device is used for guiding the waves which lie between light waves and radio waves in the electromagnetic spectrum — the microwaves. Longer than visible light waves (and invisible infra-red waves), shorter than radio waves, microwaves are becoming increasingly important as a means of communication. The waveguide takes the place of the transmission line between the oscillators and the aerial. When an alternating electric current flows along a wire, magnetic fields are set up around the wire. The quicker the oscillations, the bigger the fields. Transmission lines for carrying the alternating signal occur in pairs. The second line of the pair also has magnetic fields built up around it. Electric fields between the two wires build up and collapse. The wires are really guiding an electromagnetic wave. Some of the energy is bound to leak away (be radiated) and some is lost because the material spacing the wires is not a perfect insulator. The loss becomes more and more serious the quicker the current changes direction 1.e. the higher the frequency of the guided wave. At microwave frequencies, greater than 3,000 million cycles per second, a pair of wires becomes very inefficient as a means of carrying signals from one point to another. The two main reasons for inefficiency are (1) fields stray outside the conductor, and some of their energy is lost (2) it is not always possible to space the wires with air (the two wires must be spaced apart by some form of# A DEVICE FOR ALTERING THE INCOMING MICROWAVES i 7 A_ MICROWAVE BRANCHING ARRANGEMENT ES insulator). Other materials, less efficient than air have to be used. In waveguides, difficulties are overcome by (1) turning the two wires into a closed box, enclosing the fields and (2) there is no need for an inefficient spacer, because the sides of the waveguide are spaced rigidly apart. The space can be filled with dry air or evacuated where it joins an oscillator. How does a waveguide work? The microwaves in the guide behave almost like light waves. Ifa lamp were placed at one end of the tube, some of its light would bounce from wall to wall along the guide, criss-crossing across the airspace until it reached the far end of the waveguide. If the tube is straight, some light waves can, of course, travel straight to the far end without bouncing. This, however, cannot happen to microwaves in a guidewave. The distance apart of the waveguide walls is between a quarter and a half a microwavelength. Disturbing things happen to all kinds of waves — sound waves and water waves as well as the electromagnetic variety — whenever the guiding tube is of roughly the same dimensions as the wave. Only a few wave-paths become possible. Most The waveguide is a cross between a hollow light guide and a pair of conducting wires. 7A



WAVEGUID

A SPYGLASS tube is a guide for light waves, a pair of solid electrically-conducting wires is a guide for radio waves. The cross between the two is a hollow, electrically-conducting tube, a *waveguide*.

Not surprisingly, this hybrid device is used for guiding the waves which lie between light waves and radio waves in the electromagnetic spectrum – the *microwaves*. Longer than visible light waves (and invisible infra-red waves), shorter than radio waves, micro-

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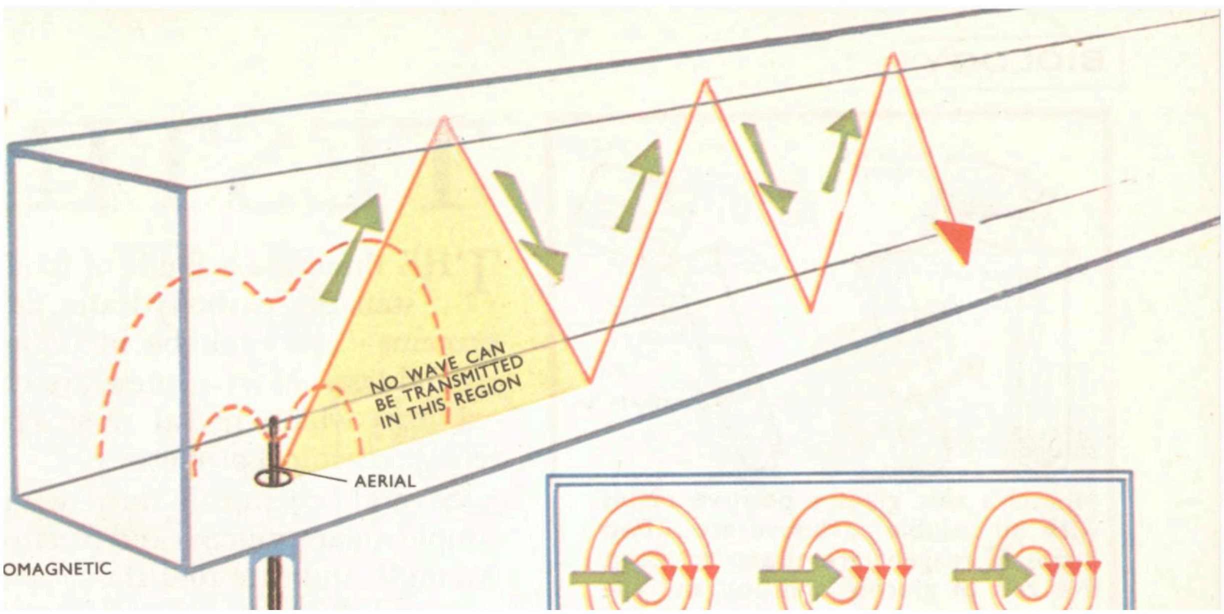
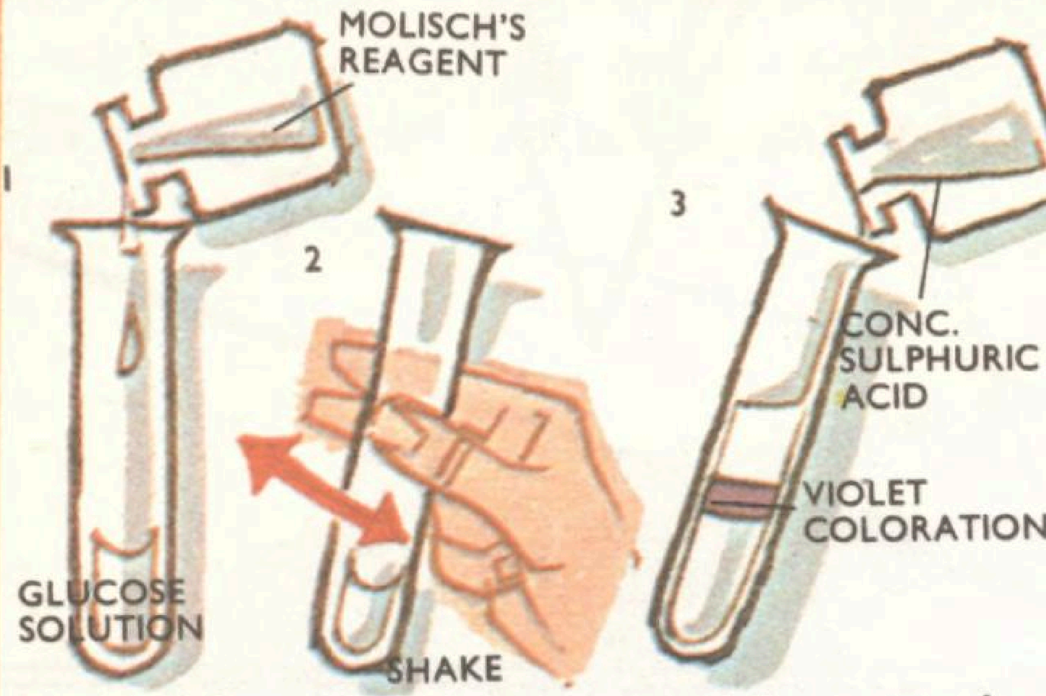
paths are ruled out because waves eliminate each other —wave crests have coincided with wave troughs, and interfered destructively. One end of the waveguide is coupled to a microwave oscillator, perhaps through a small ‘probe’ and an aerial which radiates microwaves as the lamp radiated light waves. The aerial radiates in nearly all directions, but only waves radiated at certain angles, relative to the waveguide, can crisscross successfully to the other end. Each time the wave ‘strikes’ the wall, magnetic and electric fields are set up in and around the wall, and electric currents are induced to flow to-and-fro along the waveguide. Current flows along the two of the sides of the guide (corresponding to the two wires of transmission lines), changing direction as the microwave changes direction, i.e. at the microwave frequency. Certain conditions must be fulfilled. The wave is surrounded on all sides by a conducting metal tube. Lines of electrical field joining temporary accumulations of positive charge to temporary accumulations of negative charge must always be at right angles to the inside surface of the guide. The magnetic field, on the other hand, must lie parallel to the surface. Only a few waves can satisfy these conditions for a given waveguide. The wavelength of the wave is related to the diameter of the waveguide, and also to its length. The smaller the microwave, the smaller the tube diameter. A 10 cm wavelength microwave could be propagated along a rectangular ELECTROMAGNETIC WAVES One way of introducing waves to a waveguide. The wave front bounces between the walls of the guide. waveguide of inside dimensions 2-84 x 1:34 cms. The waveguide for I cm microwaves would be almost exactly 10 times smaller. Most waveguides have a rectangular cross section, but circular tubes are Boundary Condition (1) _ Magnetic Field MAGNETIC FIELD WAVEGUIDE Magnetic fields must be parallel to conducting surfaces. sometimes used. The waveguide is rarely a straight, uninterrupted tube. With care, the tubes can be bent to guide the wave round corners. When different circuits join on to one waveguide, extra rectangular and circular pieces are joined on. These can act as a short circuit (a shunt), a complete wave-path blockage, a high-power/ low-power switch, and so on. Whenever two pieces of guide af€ joined™ é — together, cavities specially-shaped have to be made around the join so that the waves do not leak away. Somewhere near the end of a waveguide is a slot, or a series of slots half a wavelength long. These are cut into Boundary Condition (2)

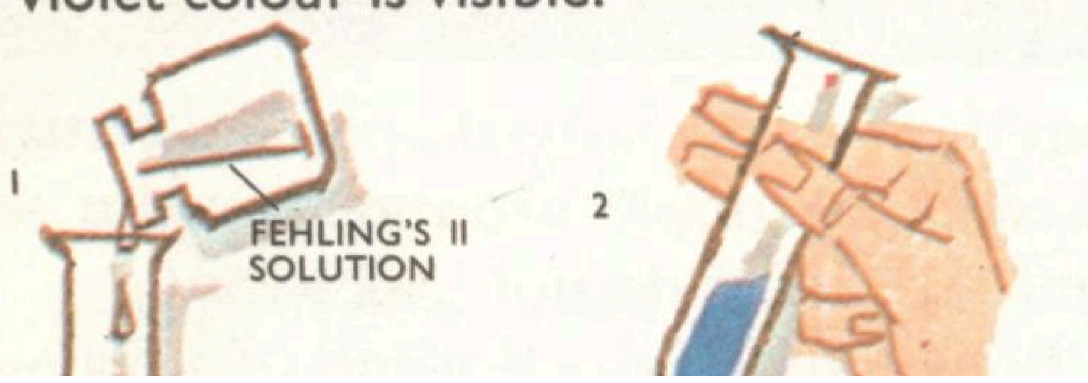
WAVEGUIDES Electric fields always lie perpendicular to conducting surfaces. the waveguide, and positioned so that they act as aerials to radiate the wave away from the guide (waveguides can be used in reverse — the slots then act as aerials for incoming microwaves). The slots are cut so that they disturb the waves in the guide as much as possible. Diagonally-placed and off-centred slots, half a signal wavelength long, interrupt the field pattern, and radiate signal energy IPN PATTERNS REINFORCE Pd EACH OTHER } TWO WAVE . 2063

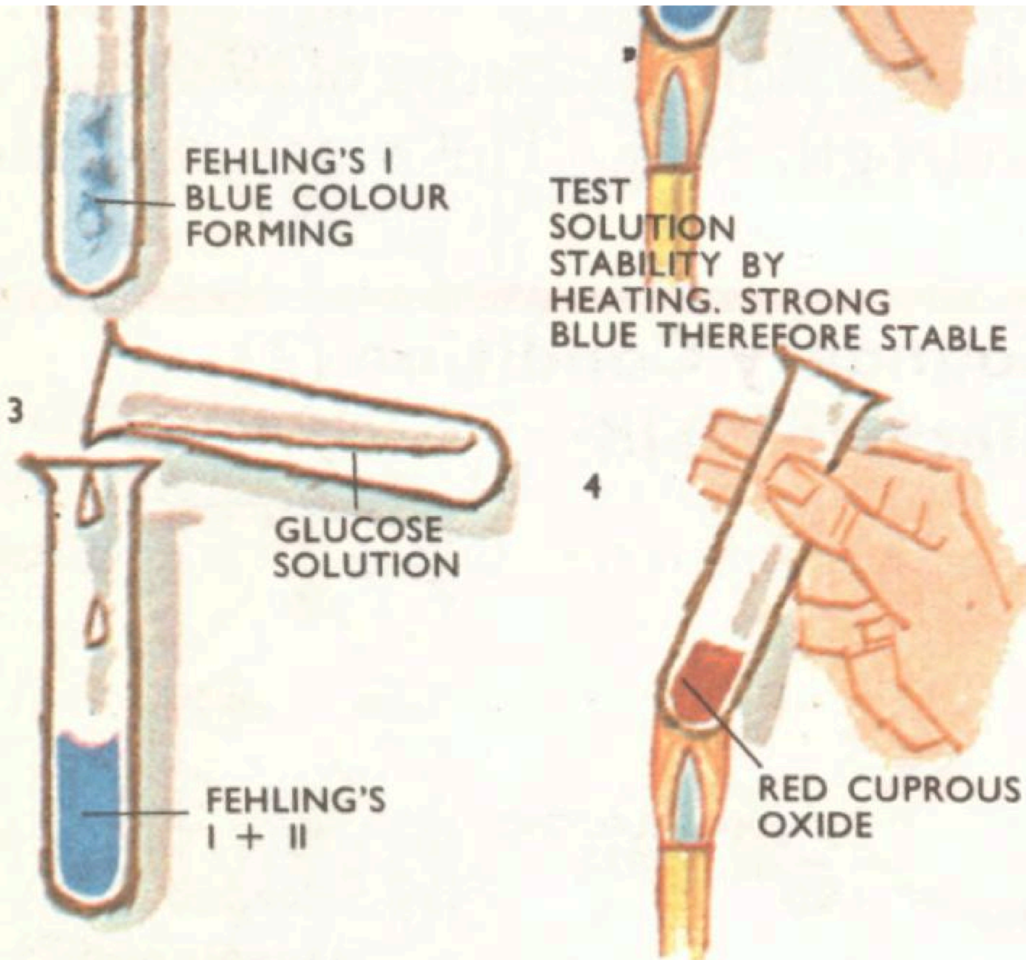


BIOLOGY Molisch’ s test gives a caitive result with all soluble carbohydrates. Two drops of reagent are added to about five cc’s of glucose solution and the mixture shaken. An equal quantity of conc. sulphuric acid is poured slowly down the side of the inclined tube to avoid mixing. At the boundary between the two solutions a reddishviolet colour is visible. i FEHLING'S II _ SOLUTION ~ FEHLING’S | BLUE COLOUR _ FORMING LI B HEATING. STRONG BLUE THEREFORE STABLE 3 GLUCOSE — SOLUTION HLING’S \} RED CUPROUS Hier 1“ OXIDE Sugars that are easily oxidized (i.e. reducing sugars) reduce copper solutions with the formation of red or yellow cuprous oxide depending on the experimental conditions (speed of reduction etc). Typical reagents are Fehling’s solution, Barfoed’s solution and Benedict’s solution. TEST FOR STARCH _ TUBE BLUE STAR COOLS IODIDE Test for Starch. The addition of a few drops of iodine solution to starch solution produces a characteristic dark blue starch iodide. Starch may easily be extracted from potato or a similar materia! rich in starch. 2064 TESTING FOODS HE three main kinds of food substances — carbohydrates, fats and proteins — can each be identified by special tests. Most of these are chemical tests which reveal their characteristic chemical properties. Several chemical tests with the’ simple sugars, glucose and fructose, for example, indicate that they are easily oxidized and have reducing properties similar to those of aldehydes and ketones. Fehling’s solution and Benedict’s solution both yield characteristic precipitates, the colour varying with the concentration of sugar. Similarly, ammoniacal silver nitrate is reduced producing a silver mirror on the inside of the test tube. Further study has in fact revealed that glucose contains the —CHO grouping characteristic of aldehydes whilst fructose contains the =CO group of ketones (glucose is consequently referred to as an aldose, fructose is a ketose). Some of the dissaccharides (e.g. maltose) are also easily oxidized and are ealied reducing sugars, others, such as sucrose, are not oxidized so readily and are non-reducing sugars. Other tests yield characteristic results with the more complex carbohydrates. For example, a few drops of iodine added to starch solution produce a blue/black colouration due to the formation of starch iodide. There are also identification tests for fats and proteins. If a little fat is shaken with carbon tetrachloride and a drop of the liquid placed on a clean white paper, a grease spot is obtained on the paper when all the carbon tetrachloride has evaporated. Addition of an alkali such as caustic potash produces a clear soap solution which froths readily on shaking. With the addition of the reagent Sudan III toa fat, a red colouration is formed. Another reagent, Millon's reagent, yields a red precipitate with protein. Chemical tests can be applied to a Sugars such as glucose, maltose, lactose, and arabinose form osazones with phenylhydrazine. The osazones form as yellow crystalline solids. The inset shows the characteristic appearance of the crystals of three osazones. When the sugar is present in minute quantities the crystal appearance 1s affected by the presence of impurities (Crystals of glucosazone have a rounded rather than a 'needle-shaped' appearance). : iG GLUCOSAZONE (X50) 'OSAZONE (250) LACTOSAZONE (X200)



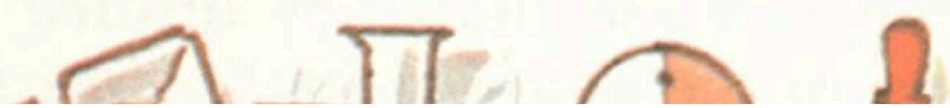
Molisch's test gives a positive result with all soluble carbohydrates. Two drops of reagent are added to about five cc's of glucose solution and the mixture shaken. An equal quantity of conc. sulphuric acid is poured slowly down the side of the inclined tube to avoid mixing. At the boundary between the two solutions a reddish-violet colour is visible.

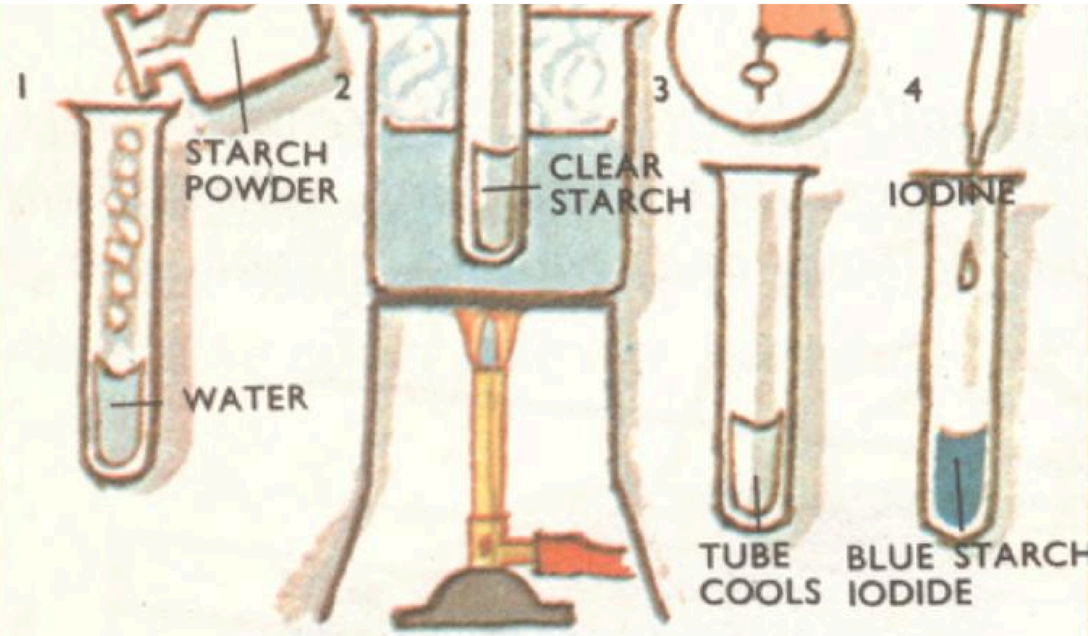




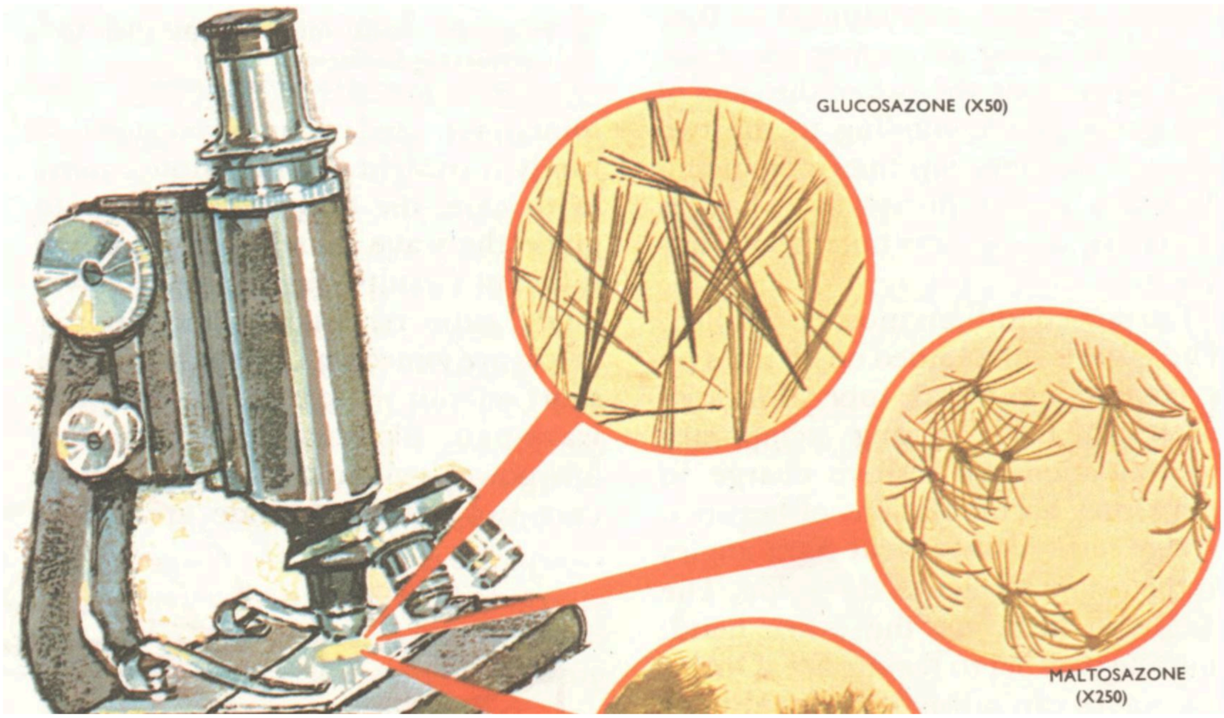
Sugars that are easily oxidized (i.e. reducing sugars) reduce copper solutions with the formation of red or yellow cuprous oxide depending on the experimental conditions (speed of reduction etc). Typical reagents are Fehling's solution, Barfoed's solution and Benedict's solution.

TEST FOR STARCH



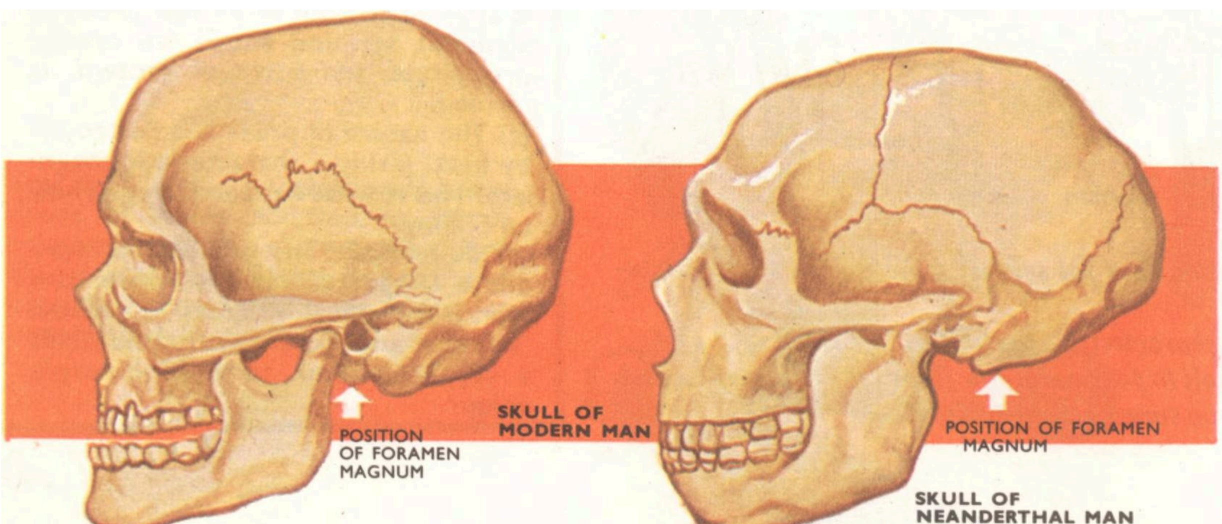


Test for Starch. The addition of a few drops of iodine solution to starch solution produces a characteristic dark blue starch iodide. Starch may easily be extracted from potato or a similar material rich in starch.


GLUCOSAZONE (X50)
MALTOSAZONE
(X250)

STARCH STARCH SOLUTION SOLUTION STARCH 7 THE DIGESTION eee BY DINE SS} SOLUTION 2 . — NO REDUCTION FEHLING'S SOLUTION FEHLING'S SALIVA SOLUTION Starch gives a negative result with Fehling's solution but after it has been digested with saliva Fehling's solution is reduced showing that the starch has been hydrolysed to give a reducing sugar. As ittle egg e boiled Millon's reagent gives a pink coagulate indicative of protein. b) Eggwhite yields a white precipitate on the addition of a few drops of concentrated nitric acid. On boiling a PROTE! ESTs 3 EGG (A) | 2 MILLON'S REAGENT RED PPT. BLUE COLOUR REMAINS AMMONIA SOLUTION wide variety of foods — egg yolk, milk, | crushing seeds, plant storage organs raisins, potatoes, onions, peas, beans, _ and the like in suitable solutions, the ground-up hazelnuts etc.—and the food present can be extracted and YELLOW food substances within them identified. tested and so certain aspects of plant oe Their value as foods can therefore be | metabolism investigated. ascertained to a certain extent. By RANGE FAT TESTS BUTTER Ww, DILUTE COPPER SULPHATE CARBON TETRACHL yellow colour results and if the mixture is allowed to cool and a few drops of ammonia solution added, an orange precipitate, indicative of protein, is | formed. _ (c) The nature of protein is destroyed by heat, giving a characteristic coagulate. This is what happens to eggwhite on boiling. (d) Buiret test. The addition of very dilute copper sulphate solution to a mixture of equal volumes of protein 'solution' and sodium hydroxide yields a very pale pink to reddish-purple colour. SUDAN Ii! 2 (c) SOAPY JESOLUTION RED OIL PRECIPITATE (a) Fats give a grease spot if a drop of solution in a fat soluble material, such as carbontetrachloride, 1s allowed to evaporate. (b) 2 or 3 drops of Sudar III reagent yield a characteristic precipitate. (c) The addition of caustic potash ov other alkali to fat produces a clear soapy solution which froths on shaking. Soaps are the sodium or potassium salts of fatty acids. 2065

ANTHROPOLOGY Homo Neanderthalensis— Man’s Extinct Gousin NEANDERTHAL man typifies the popular conception of primitive man — rugged and somewhat ape-like, yet sufficiently human to fully justify the term ‘man’. He takes his name from the Neander Valley near Dusseldorf in Germany, where the first properly-studied remains were found in 1856. These remains — consisting of a piece of skull and a few other bones — were the first to be recognized as fossil man, but other, earlier-found, fossils that were not studied at the time have now been shown to be remains of Neanderthal man. The Dusseldorf remains showed similarities with both man and apes and anthropologists regarded Neanderthal man as a stage in the evolution of modern man from apes. Thomas Huxley, in his book ‘Man’s Place in Nature’, compared the apes and Man and showed how the then recently-discovered fossils fitted into the proposed pattern of evolution. If the Dusseldorf fossils alone had come to light, we might still have regarded Homo neanderthalensis as our direct ancestor. But many more fossils have been found, thanks largely to the Neanderthals’ habit of burying their dead, and the theory has had to be altered to fit the facts. Remains of more than 50 individual Neanderthals have been found, some of them almost complete skeletons — like the ‘old man of La Chapelle’, a nearly perfect skeleton of an old man found in a cave in the Dordogne region of France. All of the Neanderthal fossils date from the third interglacial period or the early part of the last ice advance, giving them an antiquity of between 100 and 200 thousand years. The fact that most of the skeletons > / have been found in caves has led to the inevitable necessarily so, however, for it may be that they just preferred to bury their dead in caves, or that remains are a preserved better in caves. Archaeologists, too, concentrate on cave depe and so it follows that most of th finds come from caves. Neand@at 2 Man certainly entered cayé# jamg probably lived in them at tim@g@i there is no proof that he always lived in them. Examination of the skeletal remains and associated tools or artefacts has enabled scientists to build up a fairly complete picture of what Neanderthal man was like and what sort of life he led. The typical Neanderthal barely topped five feet in height but the limb bones were thick and heavy, indicating a powerful and muscular body rather like that of the modern apes. He is belief that Neanderthal |S Man was a cave dweller. This is not Skull of Neanderthal Man compared with skull of modern Man. The foramen magnum was Sarther back than in modern Man so that the skull would have been carried in a more forward position. The brow ridges were far more massive and the skull bones generally thicker. OF FORAMEN MAGNUM SKULL OF NEANDERTHAL MAN frequently pictured in a_ stooping attitude but the evidence for this is not conclusive and the stoop has probably been exaggerated. The bones of the lower leg suggest that he often squatted on his haunches as many children do today. The skull must have been carried in a more forward position than that of modern Man, for the foramen magnum (the opening where the vertebral column articulates with the skull and through which the spinal cord passes) was farther back than in modern Man. Confirmation of this posture is provided by the neck vertebrae which have very long ‘spines’, presumably to anchor the large muscles necessary to balance the head in its forward position. The skull as a whole was large, with an average brain size of about 1,450 cubic centimetres (millilitres) — compared with 1,350 cc for the average modern human brain. The shape of the skull, however, showed certain primitive features. The skull bones were thick and there were relatively



huge brow ridges about the eye sockets. There was not much of a fore-head, the brain case sloping gradually back from the eyes. The jaws protruded and the lower one sloped backwards from the mouth without forming a chin. The shape and arrangement of the teeth was essentially human but they were of a larger size than that generally found in modern Man. During the life of an individual, the bone surfaces and joints undergo changes and the amount of change, such as the closing up of the skull joints, is a good guide to the age of the individual. The state of the bones of the Neanderthal fossils indicates that few of them lived to a good age — even the so-called 'old man of La Chapelle' died at about forty. Lack of hygiene must have resulted in a lot of disease which, together with the danger from wild animals and, especially in the colder periods, starvation, probably resulted in early death. Female skeletons are more rarely found and few Stone Age women seem to have even reached the age of thirty. ie gee a Some activities of Neanderthal Man. Top left, cleaning an animal skin with a stone scraper. Right, hunting wild animals provided Neanderthal Man with food. Bottom, fashioning stone tools and hardening the points of wooden spears by fire. Tool-making among the Neanderthals The Neanderthals used quite a wide range of stone tools, fashioned mainly from flint. The tools were sufficiently different from those of earlier periods for a separate culture period to be established. Typical collections of tools were found in a cave at Le Moustier in the Dordogne, and the name 'Mousterian' has been given to this period of stone age culture. Neanderthal Man has, accordingly, also been called Mousterian Man. Most of the tools were made from flakes of flint by carefully chipping the edges. This produced scraping and cutting tools for use in skinning and cutting up the animals that had been killed for food and skins. Some of the early Mousterian hand-axes bear signs of having been derived from the earlier Acheulian traditions. Neanderthal Man used fire regularly and it probably served several purposes — to keep him warm, to drive away wolves and other carnivores, and to sharpen the points of the hunters' wooden spears. Broken bones seem to have been used for tools and weapons but there is no evidence that the Neanderthals had mastered the art of shaping bone. Hunting and Social Life Other bones associated with the Neanderthal remains indicate the sort of food on which the Neanderthals existed. Many of the animals that they hunted were very large and it is obvious that Homo neanderthalensis must

have worked in groups to capture these animals. Wild boar, ibex, cave bear, rhinoceros, and even elephant figured in the diet of the Neanderthals, according to the animal bones found with the human remains. Naturally, not all these animals were hunted at one period. Changing climates brought various animals within reach of the Neanderthals. The animals were probably killed largely with the wooden spears that had been sharpened in the fire, but the occurrence of some large battered skulls suggests another possible method of hunting. These skulls, notably those of the cave bear, have been found in association with some large boulders and it is likely that the Neanderthals stood high above the bears' runs and rolled the boulders down onto the animals. Larger creatures, such as Complete skeletons of modern Man and Neanderthal Man compared. Neanderthal Man, carrying his head in a more forward position, did not walk fully upright. His stoop, however, has often been exaggerated. Neanderthal skeletons are common in Europe but similar forms have also been found in South Africa and Fava.
2067

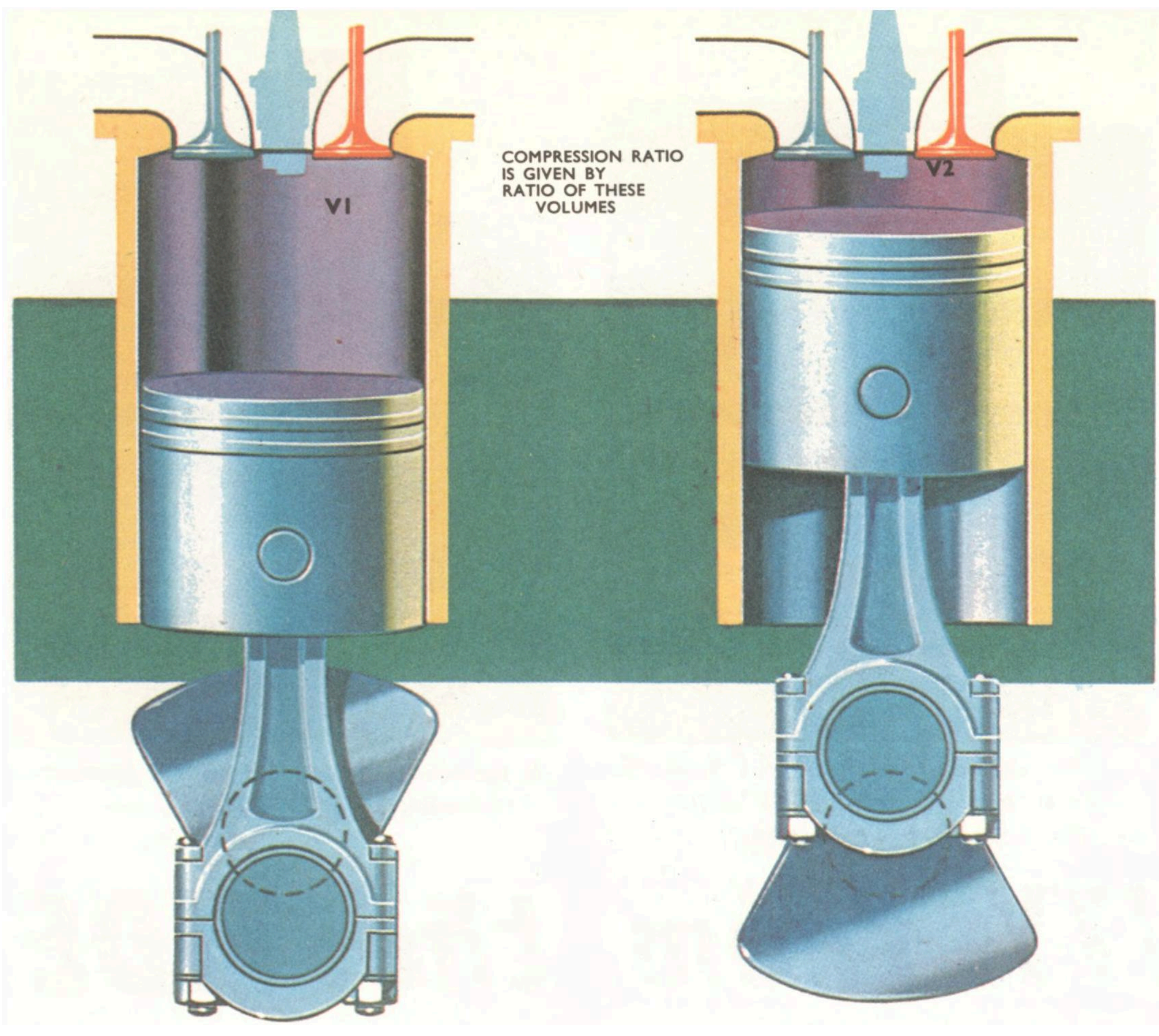
Some activities of Neanderthal Man. Top left, cleaning an animal skin with a stone

MODERN MAN PRESENT-DAY THEORY Left, old idea of Neanderthal Man's relationship with modern Man. Neanderthal Man was considered an intermediate stage in the evolution from Pithecanthropus to modern Man. Right, today's generally-held theory — Neanderthal Man probably evolved in a line parallel to that of modern Man. Both lines spring from a common stock so that Neanderthal Man becomes Man's cousin — not his direct ancestor. the rhino and elephant, were possibly killed in pit traps. Neanderthal Man probably worked in family groups of twenty or more individuals and it is highly unlikely that they could have done this without at least some sort of language. Although the brain was of a simple form compared with that of modern Man, the regions corresponding to that concerning speech in Homo sapiens were probably well developed. It can be assumed that Neanderthal Man could speak though whether he reached a very high level of articulation is not certain. Distribution of Neanderthals The remains of typical Neanderthals are found over large areas of Europe and neighbouring parts of Asia and North Africa but the Neanderthaloid type was by no means confined to these areas. Rhodesian Man, whose skull and other ones were found at Broken Hill in Northern Rhodesia in 1921, is clearly related to Neanderthal Man, although there are some differences in the structure of the limb bones. The geologica! evidence indicates that Rhodesian Man lived at a later date 2068 than the typical Neanderthals. The associated stone tools are of a more primitive type than the Mousterian implements of Europe, but this is of no great significance, for cultures of this sort take some time to spread and a primitive tool may still be in use in one area centuries after it had been superseded in others. Skulls of Solo Man — found in late Ice Age deposits on the banks of the Solo River in Java — also show numerous resemblances to the Neanderthals, but they are not complete enough to make detailed comparisons. It is likely that both Rhodesian and Solo Men were merely variants of the Neanderthal stock, and like it, became extinct. The Relationships of Neanderthal Man The previous paragraph showed that men of the Neanderthal types were widespread during the third interglacial and the early part of the last ice advance. Fossils from an even earlier period show a number of similarities with both Neanderthals and the older Pithecanthropus of the midPleistocene. It seems that Pithecanthropus gave rise to the generalized Neanderthal type during the latter part of the Pleistocene. When the first

Neanderthal fossils were studied it was assumed that the creature was a stage in the evolution of Modern Man, and, if further discoveries had not been made, it is likely that we would still hold this view, for the Dusseldorf fossils certainly show a mixture of apelike and human characters. We now believe, however, that the Neanderthals were no more than cousins of modern Man and there is plenty of evidence to support this. The earliest Neanderthal fossils known are only about 200,000 years old, yet, before this, there were creatures, such as Swanscombe Man, that showed modern characteristics. The brain of the typical Neanderthal was larger than that of modern Man and it is highly unlikely that it would have become smaller again during evolution. Another very important point is that the later Neanderthals became progressively more heavily built — the jaws and brow-ridges developed and the creatures become more and more unlike Homo sapiens. This is quite the reverse of what would be expected if the Neanderthals were the ancestors of modern Man. Finally, the Neanderthals were replaced in Europe, very suddenly by a modern type of Man. There was no gradual change such as would have occurred if one had evolved into the other. The Mousterian tools were succeeded suddenly in the deposits by Aurignacian tools of Cro-Magnon Man who must have reached Europe from elsewhere and killed or driven out the Neanderthals. Neanderthals in other parts of the world were probably replaced soon afterwards and, as far as his physical characters were concerned, Man's evolution was complete. It is believed that the Pithecanthropus type gave rise to the Neanderthal type but that, before the typical Neanderthal characteristics had evolved, the line split into two. One branch led to the extreme Neanderthals and the other to Homo sapiens so that the typical Neanderthal cannot be regarded as a direct ancestor of Modern Man. He belonged to a side-line of evolution which died out without leaving any descendants.

COMPRESSION RATIO IS GIVEN BY RATIO OF THESE VOLUMES Compression Ratios SINCE its earliest days, the motor car has been gradually improved in design, so even the cheapest of present day models is capable of operating efficiently with low petrol consumption. One of the developments that has taken place over the years is the gradual adoption of engines of higher and higher compression ratio. In the internal combustion engine the ignition of a mixture of air and petrol vapour causes a sudden expansion of the gas volume in the cylinder. This causes the piston to be ‘thrown’ into its expansion stroke and the movement is eventually translated into the forward movement of the car. A small volume of gas has been converted into a big volume of gas, and the change in volume is a measure of the effectiveness of the engine. The ratio final gas volume enclosed initial gas volume enclosed is called the compression ratio of the engine. In general, the higher the compression ratio, the more efficient the engine and the lower the petrol consumption. To increase the compression ratio is simple enough, and in fact, whereas the compression ratios in older models of mass-produced cars were only 5 or 5:5, many modern models have engines of compression ratio 8-5. But there is a practical limit to the value of compression ratio that can be used. This is imposed by the behaviour of the air-petrol mixture — in particular its tendency to knocking or pinking under high compression. When an engine is working properly the ignition of the mixture should result in a flame starting at the sparking plug and steadily spreading throughout the gas volume. If the compression ratio is too high, a part of the gas volume, ahead of the flame front, becomes overheated and over-compressed and explodes spontaneously causing a loud knocking noise. This results in inefficient and laboured operation of the engine, because the piston is receiving an extra ‘push’ at the wrong time (it is rather like a child’s swing being pushed when it is not at its highest point of swing). This occurs only when certain types of petrol are used. Knocking will not occur if a petrol of sufficiently high TECHNOLOGY The compression ratio of an engine is the ratio of the volumes of atr and petrol vapour enclosed in the cylinder at the bottom of the stroke (Vr) and the top of the stroke (V2). octane value is used. Experimental and standardizing fuels can be blended to have a particular octane value. For example, pure heptane (C7H4¢) has a very poor resistance to knocking, but pure iso-octane (CgHis) has a high resistance.

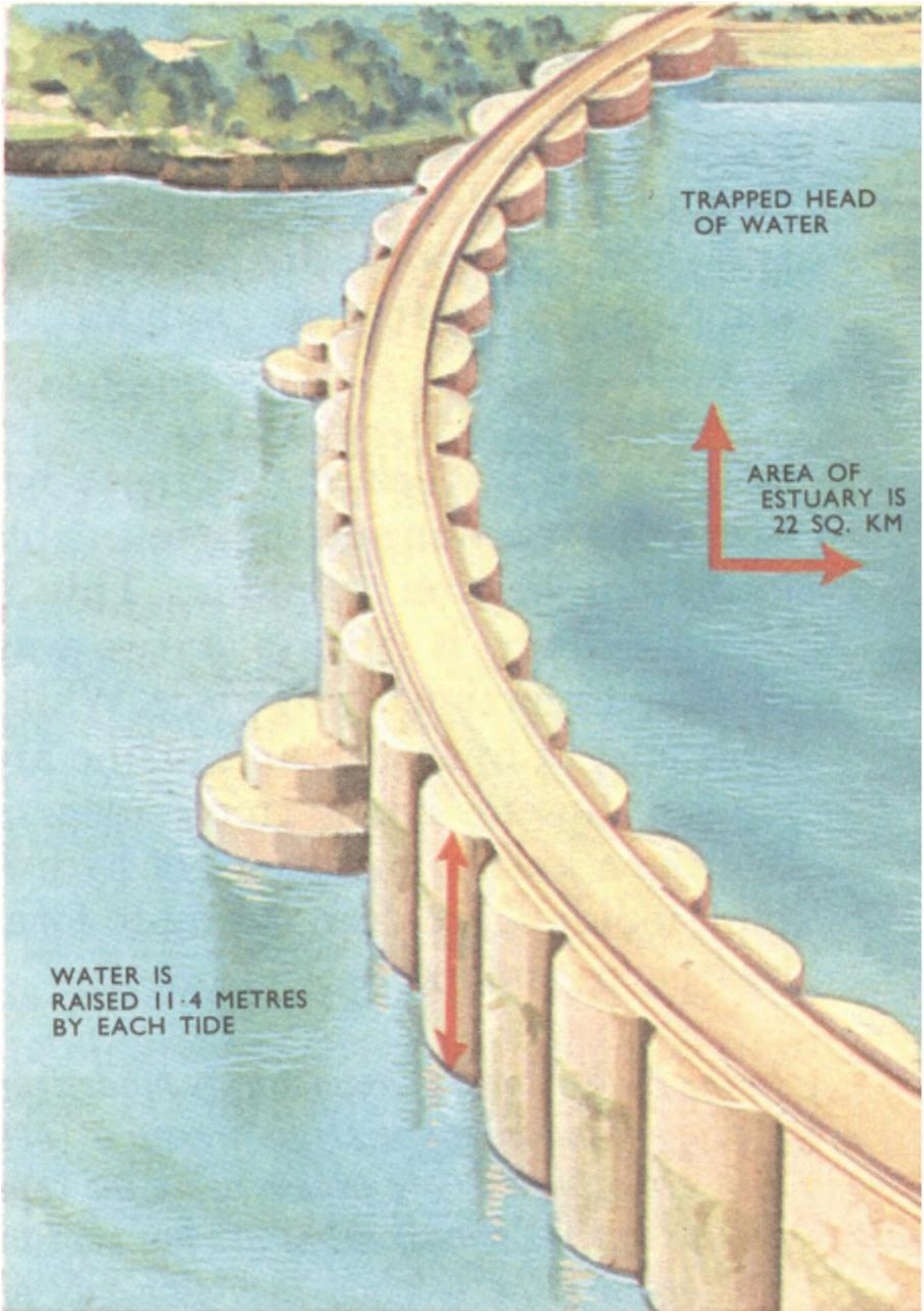
Motor fuels are tested and rated by blending these two liquids and using the behaviour of the mixture as a basis for comparison. A special test engine is used, in which the compression ratio can be varied, and compression ratio is found (for the fuel under test) at which knocking occurs. The same engine is then run on a mixture of heptane and octane and the constitution which just produces knocking is found. For example, it might be found that a mixture containing 75% octane first produces knocking at that particular compression ratio. The petrol is then said to have an octane rating value of 75. A 100 octane petrol has the same resistance to knocking as 100% pure isooctane. Resistance to knocking can be improved by the use of certain additives, most usually tetra ethyl lead. If the compression ratio is too high for the particular petrol knocking or pinking occurs. Some of the mixture ignites ahead of the flame causing a spurious explosion.



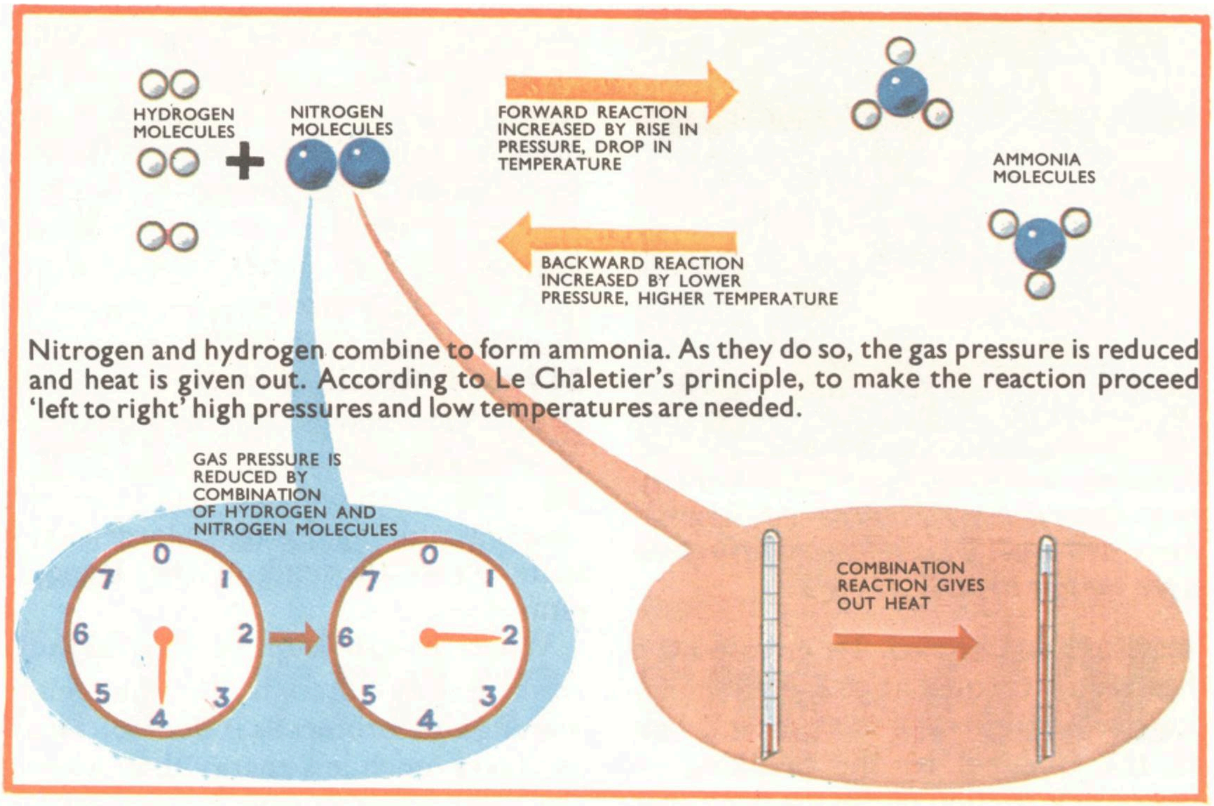
2 —— Oe ee eT Turbine turns to supply a little electricity to the system. ELECTRIC AN has so far done very little to capture the energy of the seas and put it to work for him. The French Tidal Power Station of the Rance, when it is completed, will have the distinction of being the only one of its kind in the World to produce electricity from the regular ebbing and flowing of the tides. It so happens that on this particular stretch of French coast, the difference Sources of Electrical Energy By far the largest proportion of electrical energy produced at present comes from fuels, in one form or other, _ extracted from the Earth. These fuels — coal, oil, and uranium all have to be extracted, refined and transported _ before they can be used to raise water into steam that drives the electrical generators. In addition to these — sources of energy, there are other natural sources — fast flowing rivers and heat from the Sun. ¥ In all the conventional methods the energy locked up in a fuel is first converted into heat energy. In coal an oil it is chemical energy that is converted, in uranium the energy released by controlled fission of the uraniu' nuclei is used. In the Rance scheme, the mechanical energy of the tide is directly converted into electrical energy. This is . similar, in principle, to hydroelectric schemes in operation all over the world, where water delivers its mechani- HYDROELECTRIC POURS. WATER E IC BORROWED THROUGH FROM SYSTEM TO' DRIVE TURBINE A little electricity ts borrowed from the system to turn the turbines to artificially raise the water level in the estuary. from between high and low water levels is sufficiently great to be able to run an efficient power plant. There are, in fact, very few parts of the globe where the water level rises and falls enough to make this a worthwhile venture. Large scale operations were once started at the Canadian/ American border. Five thousand men were working on the project but it had to be abandoned because of the high SMALL AMOUNT OF LECTRICITY As the basin empties, the turbine produces a lot of electricity for the system. the TIDE cost of transporting the cable to where it was needed. The difference between high water and low water in the Rance Estuary is on average 11-4 metres and there is also an immediate use for the electricity produced in the area to make the scheme practical. Twice every lunar day (24 hours 50 minutes) an 'astronomical wave' comes from the open Atlantic and sweeps into CONVENTIONAL POWER STATION NUCLEAR POWER STATION cal energy as it falls from one level to another. This energy is

directly converted into electrical energy. The POWER TURBINE Rance scheme is also very much like a number of pump-storage schemes used in Luxembourg at Vianden, andin = a a Wales at Ffestiniog. In both of these projects, as in the Rance one, water is not allowed to flow through the a 4 generating sets until it is required, at periods demanded in the electrical supply system. In the pump-storage schemes water is pumped up a hill, from a lower reservoir to a higher one, at a time of day when there is not a great demand for electricity. It is then held in reserve until the demand is at a peak and is then released and allowed to flow through the generating sets and produce much-needed extra electrical power. Another example of conversion of natural energy is in the use of solar energy. A vast amount of radiant energy from the Sun is pouring down on the surface of the Earth all the time and this can be used to raise water . — pie steam can then drive generator turbines. Such schemes have been put into operation in parts of the Soviet Union. :

When the basin is empty, a little electricity is borrowed to turn the turbine and empty the basin even more. This is handsomely repaid when the tide turns. the English Channel. Its gross power has been estimated at 56 million horse power. About half of this power is lost in the Channel by the breaking of waves, the friction of the sea floor and the splashing of water along the length of the coast. It is part of this lost energy that the engineers hope to capture with their tidal power stations. The principle of the whole operation is very simple. Man has been using it since antiquity with tidal mill wheels. A creek with a sluice gate opened to the incoming tide will fill with water. This can then be trapped by closing the gate and the trapped head of water later used to do work or run some sort of generating plant when the tide has gone out. Choosing a Site The basic power of a tidal power station depends on two things, the area of the inlet or creek in which the water is trapped and the difference between high and low tide. It is therefore important to choose a site where this value is as large as possible. The Rance estuary has an area of 22 square kilometres and the water level rises || -4 metres on average each tide which means that a slice of water 184 million cubic metres flows in and out twice a day. The Mediterranean could never rely on the tide for its power as the tide does not rise and fall more than a few decimetres. Unfortunately, this very simple theory would fall down in practice for it means that electricity can only be generated when the tide is on its way out and a high sporadic output of electricity, if it occurs in the middle of the night, is no use to anyone. A regular output at the right time is needed, calling for a much more sophisticated arrangement. In fact, to arrange the supply to meet the demand requires a power station computer to direct the operations of opening and shutting the sluice gates. 24 Power Stations in One The Rance power station organizes its power output by means of 24 bulb sets which to an onlooker look like 24 channels running through a large dam built across the end of the Rance estuary. When the tide is on its way in, the water rushes through the channels, turning the turbines as it does so. This produces much less energy than when the basins are emptying, but nevertheless, it is still worthwhile. At the end of the tide, energy is borrowed from the network so that the turbine continues to turn, overfilling the basin behind the dam in the process. Overfilling is a way of getting something for nothing. It costs very little to take electricity from the system to artificially raise the water level by a further 50 centimetres. But three hours later, when the surrounding outflowing tide has fallen by some 6 metres this same water will have a head of 6 metres and will in practice be able to provide nearly 12 times as much energy as that expended in raising it in the first place. The amount of time the water is stored at its artificial level depends on the demand for the electricity it is capable of producing. At the right time, the water is allowed to flow out to sea while the turbines supply electricity to the network. The turbine which converts the flow of water into useful electrical current is a sunken submarine right in the path of the water current, surrounded by water on all sides and possessing a large propeller which is moved by the current. The propeller has adjustable blades which can be arranged to suit the conditions and the sunken submarine can be reached by means of a shaft with a ladder. When a basin has emptied and reached the level of the tide around it, a little energy is again borrowed from the network only this time to produce over-emptying. The whole cycle can then start all over again. Each little power station is operated quite independently to suit the different tides and output needed. The number of different cycles which can be utilized in the course of one month is staggering and makes it possible to deal with the most varied rates of output and allow for all tidal possibilities. This must take into account the fact Twice a day 184 million cubic metres of water rushes through the dam dropping through a distance of 11-4 metres and thus providing power. The times of day are selected to provide power when it is most needed — at times of peak demand. that the tides occur 50 minutes later every day and do not rise by the same amount each time. It is estimated that the annual output of the station will be around 540 million kilowatt hours per year, a very small output for a power station. But it must be remembered that the fuel will never run out and the plant will act as a useful indicator for deciding whether the same principle can be used elsewhere. 2071


TRAPPED HEAD
OF WATER
AREA OF
ESTUARY IS
22 SQ. KM
WATER IS
RAISED 11·4 METRES
BY EACH TIDE

PHYSICAL CHEMISTRY Le Chatelier's Principle HEN acake is baked, an irreversible process has taken place. The ingredients — flour, fat, sugar, baking powder and currants — have been mixed together and heated and it is then virtually impossible to separate them out again. In chemistry too, there are many irreversible processes. If a mixture of iron filings and flowers of sulphur is heated, the chemical compound iron sulphide is formed. This process cannot be reversed to allow the iron and the sulphur to be separated. But there are also a large number of reversible processes in chemistry. For example, if volumes of hydrogen gas and iodine vapour are mixed the compound, hydrogen iodide, is formed: H, + I, — 2HI There is an important difference between this reaction and the irreversible formation of iron sulphide. This is because there is always a two-way reversible reaction taking place. Whilst molecules of hydrogen and iodine are joining up to create molecules of hydrogen iodide, molecules of hydrogen iodide are dissociating to form hydrogen and iodine: In the Haber process ammonia is made from hydrogen and nitrogen.High pressures (over 100 atmospheres) are produced by the compressor but the temperatures are not kept too low because the reaction then proceeds too slowly. COMPRESSOR * +,° 4 ° o,¢, COOLANT ENTERS LIQUID AMMONIA 2072 OUT NITROGEN MOLECULES HYDROGEN MOLECULES ele) ee) Nitrogen and hydrog: and heat is given ou 'left to right' high pre GAS PRESS REDUCED B COMBINATIC 2HI — I, + He The reversible reaction is written : H, + I, = 2HI The proportions of hydrogen, iodine, and hydrogen iodide finally reached in the reaction (i.e. at equilibrium) can be changed by altering the temperature or gas pressure, or the concentration of any of three reagents. How is it possible to predict the way these factors will alter the final proportions reached ? It is obviously useful to know the answer because it will affect the yield in, for example, an industrial chemical reaction. The answer comes from a principle stated by Henry Le Chatelier in 1888. He stated that if there is a change in an external factor affecting a reversible reaction, the reaction will adjust itself in such a way as to counteract the change. In the reversible formation of ammonia from nitrogen and hydrogen, four molecules (one of nitrogen, three of hydrogen) combine to make two of ammonia. Furthermore, heat is given out in this reaction: No af 3H, = 2NHs + heat What is the effect on this reaction of changing gas pressure? When ammonia is formed, there is a reduction in the number of gas molecules, and the gas pressure is reduced. So, according FORWARD REACTION INCREASED BY RISE IN PRESSURE, DROP IN TEMPERATURE ‘ BACKWARD REACTION INCREASED BY LOWER PRESSURE, HIGHER TEMPERATURE Bo AMMONIA MOLECULES a form ammonia. As they do so, the gas pressure is reduced e Chaletier’s principle, to make the reaction proceed eratures are needed. ———— ED to Le Chatelier, an increase in external gas pressure will be counteracted by an increase in the number of ammonia molecules formed because this will counteract the effect of the increasing gas pressure. Increase in gas pressure therefore increases ammonia yield. The yield is affected by a change in temperature, too. According to Le Chatelier, a reduced temperature should increase ammonia yield because the reaction will proceed in such a way as to give off more heat, to counteract the drop in temperature. This is accompanied by an increase in ammonia formation. But there is a complication here, because although the final proportion of ammonia present is increased, the drop in temperature reduces the rate of the reaction, so the result takes longer to be achieved. For this reason, the temperature in the industrial production of ammonia is not kept much below 500°C. Lastly, the change in concentration of the reagent has an effect on the final equilibrium yield. If, for example, nitrogen is added, Le Chatelier’s principle tells us that the reaction proceeds ‘left to right’ to counteract this, so more ammonia is formed. The same thing applies if more nitrogen is added, or ammonia removed.

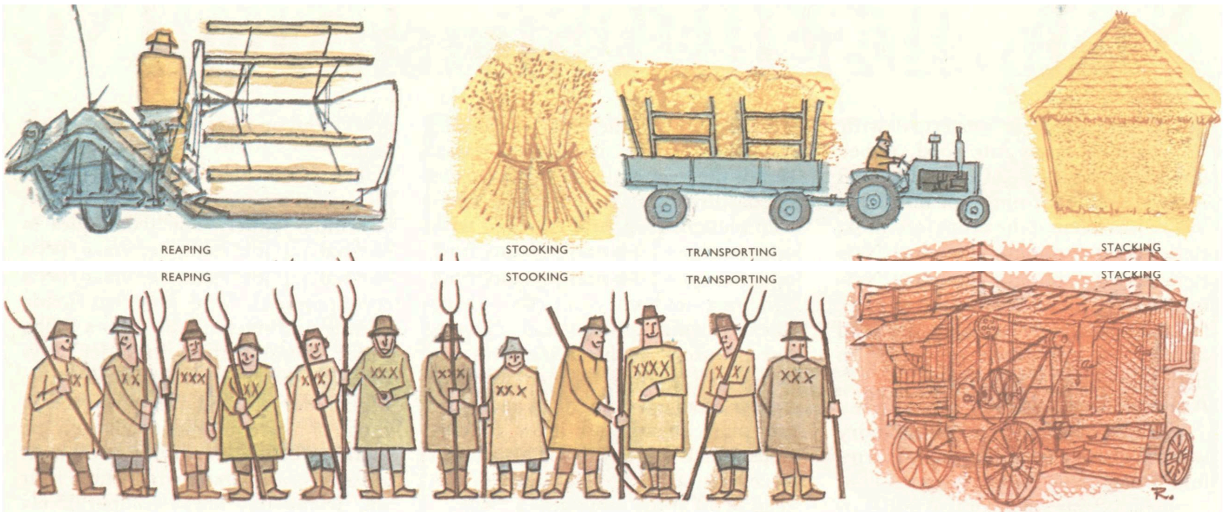


Nitrogen and hydrogen combine to form ammonia. As they do so, the gas pressure is reduced and heat is given out. According to Le Chaletier's principle, to make the reaction proceed 'left to right' high pressures and low temperatures are needed.

AGRICULTURAL SCIENCE NEW METHODS in FARMING

LL agriculture is concerned with turning crops into food, either directly as in the case of vegetables and cereals (arable farming), or indirectly by first converting the crops into milk, eggs or meat, (animalhusbandry). New methods for increasing output of food, for most efficiently utilizing the land and its resources, have proceeded in both these aspects of farming. Animal Husbandry The first development in husbandry was the change from hunting to farming, when man found it more convenient to rear and control his own animals than to prey on wild ones. Today, in an overcrowded society like our own, only an artificial system of farming can possibly meet our needs. It is necessary that every bit of land should produce as much fodder as possible and that the fodder should be converted into foodstuffs as economically as it can be. Until thirty or forty years ago the mainstay of the stock-farmer was his permanent pasture; to destroy the Section through a grain-drying silo. The increased quantity of grain produced combine harvesting, is dried and seers The silo is made of wood whtch absorbs moisture and prevents condensdtion. Nase rT DETAIL OF PRRRREERRRRRREL 7 : AUT PREREREREREERE BLOWER PIPE THROUGH CAN SERVE TWO SILOS WHICH AIR IS BLOWN precious turf of a meadow was looked upon as a crime against the land. Today the vast majority of pasture is ley pasture, that is land which has been ploughed up, and reseeded with fodder crops which may last anything from one to four years. By sowing quick-growing grasses and clovers, and by using artificial manures the farmer can produce heavier swards of greater food value; what is more, he can plan ahead and grow foddercrops which are ready for grazing in succession as they are needed, instead of all at the same time. The development of new strains has also been of great significance, especially in extending the growing season. The ‘late-bite’ of autumn grass and the spring-flush of new grass are steadily creeping round the calendar towards a winter rendezvous. This is especially important to milk producers who rely upon fresh grass to keep their output steady. New combinations like clover and rye grass or lucerne and rye grass help to grow DISTRIBUTOR ENSURES THE GRAIN IS EVENLY SPREAD AND PREVENTS POCKETS OF WET GRAIN FROM FORMING NE DISTRIBUTOR ‘ good crops on what was formerly poor land. The years since the war have also seen great advances in the production of winter fodder. The techniques of haymaking for example, have been revolutionized. The baler has made carting and stacking the crop a much easier business than in older days when it had to be handled loose. Machinery firms have produced a range of new equipment to hasten the ‘making’ of hay (i.e. drying out the cut grass). Swath-turners which turn over the drying hay are a comparatively old invention. They removed the need for gangs of workers equipped with wooden rakes. Newer machines include the ‘crimper’, which breaks or crushes the stems of grass thus helping the moisture to evaporate, and various tools designed to ‘fluff up’ the swathies (the bundles of hay) and allow air to circulate. Silage has been another modern development. Instead of drying grass out as in haymaking, it is taken while still fresh and mixed with molasses in layers in a silo or pit, the resulting ‘pickle’, though powerfully smelling, is extremely palatable — to cows. The popularity of silage has given birth to one of the newest of agricultural machines— the forage harvester. Towed by a tractor which also provides its power, the forage-harvester uses chain-like flails to cut and lacerate grass; it then blows the grass through a spout into a high-sided trailer. This machine has taken much of the labour out of silage-making. With rich ley-crops, often growing 6 inches high the ordinary process of free-grazing by cattle tends to be wasteful, much of the crop is trampled or fouled by the animals. One solution is strip grazing which is a development of the ‘folding’ system used when cattle or sheep are grazing winter fodder such as kale. Each day the animals are allowed a fresh piece of grazing, if this is of the right size it will be eaten thoroughly and the remainder of the crop remains unspoilt. Usually a portable electrified fence is 2073

REAPING STOOKING ~ TRANSPORTING It used to take 12 men 6 to 7 weeks to harvest 100 acres of corn and another 2 to 3 weeks to thresh it. COMBINING THRESHING DRYING AND STORING Now it takes 2 men using a combined harvester 2 weeks to harvest the same quantity of crop and have the dried grain safely in store. used to divide up the plots. A further development, still in the experimental stage, is ‘zero-grazing’. This system, started in the United States, goes even further and takes green fodder to the animals instead of vice versa. By using a forage harvester the crop can be gathered as it is needed, and without waste. Moreover, the animals can be given known rations which will result in the most economic foodconversion rates. This system can be used for cattle, for either milking or beef herds. In the case of beef-production much effort had gone into experiments to discover ways in which beef may be produced more quickly and cheaply. The traditional method of beef production takes two years and often involves two farmers — the rearer and the fattener ; the yearling cattle change hands as ‘stores’. The new methods speed up the process and intensify it. Instead of being reared and fattened 2074 on a basic diet of grass and hay the animals are taken as calves and fed concentrated rations based on barley —hence the term ‘barley-beef — with the addition of manufactured ‘nuts’ and just enough roughage (i.e. hay etc.) to keep the digestion healthy. This method of intensive rearing pro- duces a tender beef carcase within a year of birth — though there is criticism that it makes the meat lean and tasteless. At present, the barley-beef system is used mainly in rearing dualpurpose animals, that is calves from dairy herds, particularly Fresians, rather than the pure beef cattle which produce the top class meat. The ‘factory’ system of rearing calves for veal is based on a similar intensive system which aims at converting as much vegetable food as possible into meat. These methods, which rely on careful measurements and mixture of rations were pioneered in the pig industry where the animals are traditionally kept confined and fed on meal and concentrates rather than being allowed to forage for their food. While intensive methods of husbandry may become increasingly important in producing beef, veal and pig-meat it is unlikely that sheep will be affected ; they are in any case more efficient grazers than cattle and are often kept on marginal land, i.e. hill-pasture and poor soil, which does not lend itself to cultivation. Arable Farming One of the chief factors which has made possible the increase in cereal production has been the invention of the combine harvester. Previous methods of harvesting; cutting the corn with a reaper and binder, drying the corn in the stook, stacking the crop and threshing it later, made large labour forces necessary and placed too much reliance on an unreliable climate. At a stroke, the combine reduced harvesting to a single operation,



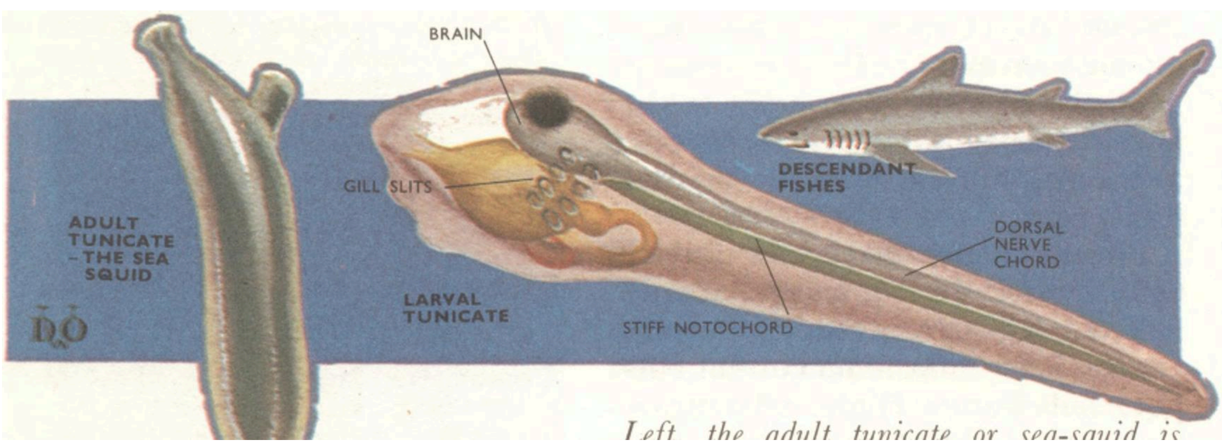
196364 1947.48 1938-39 vv Zz o) -_ < rt) = 4525 5420 8200 GALLS. GALLS. GALLS. MILK PRODUCED PER 100 ACRES Management of cattle in winter — hay, silage, kale, cattlecake, etc., are fed to cows in quantities so controlled that the minimum food outlay produces the maximum milk supply. Exact measurement means a considerable saving of fodder. The production of milk in Britain has continually risen as shown by the illustration on the left. A number of factors have played a part in the increase, from animal and plant breeding to careful measurement of food supplies allocated. only needing a few men. Continual improvement of the combine harvester made necessary a revolution in methods of handling the harvested grain. Because the corn is threshed immediately, it does not dry out as thoroughly as it did in the stook. Grain can only be stored safely if it contains less than 16% moisture; above this figure fermentation will set in, raise the temperature and eventually rot the grain. Grain drying became a necessity to cope with the larger crops which combines could handle. At the same time bulk handling of grain became increasingly important. If sacks are eliminated, and the crop is mechanically handled in bulk the saving in labour is immense. A whole new range of equipment came into use, hydraulically tipping trailers to cart the grain, cleaners to remove dirt and grass which hinder drying, augers and elevators for moving the grain, and enormous silos for storing it. The latest grain dryers combine the function of drying and bulk storage; the grain is placed in large silos in which air can be blown through it thus removing excess moisture and making it safe for storage, 50 or 60 tons to a silo. A promising new development in this field is so-called ‘wet-storage’. Grain is placed in sealed silos while still damp, the initial stages of fermentation produce carbon dioxide which replaces oxygen in the spaces between individual grains and thus prevents further deterioration. This method will obviously have important applications, especially if current experificial fertilizers put an end to this. By applying them to arable land it is now possible to grow even better cereal crops year after year without exhausting the land. At the same time the careful use of chemical weed killers has helped to keep crops ‘clean’. It is. certain that in the future arable farming will rely more and more on The development of root crops revolutionized animal husbandry 300 years ago enabling cattle to be kept through the winter months. The revolution has proceeded further, silage 1s made today and haymaking ts much faster and more efficient using machinery. ments are successful in proving that it has no adverse effect on germination in seed corn. The second major factor has been the development of artificial fertilizers which has largely put an end to the rotational system of farming in corn-growing areas. For 200 years arable farming has been dominated by the need to intersperse grain crop with greens and root crops in order to rest the land and restore nitrogen and other elements absorbed by the growing corn. The only alternative was very heavy use of animal manure which was not practical on a very large scale. The development of artiPlant breeding ts lengthening the ‘season’ of grass crops, and consequently decreasing the quantity of winter fodder needed. Thus strains of autumn grasses last much longer, and spring grasses shoot earlier. science — to analyse and repair deficiencies in the soil, to provide artificial fertilizers, chemical pesticides and herbicides. To breed more prolific and disease-resistant strains of the main cereal crops. As well as this, the engineers will produce better machines to tackle bigger jobs more quickly, so cutting down the labour required and making the farmer more and more independent of the climate.

ELECTRONICS TUNNEL DIODES "TUNNEL diodes can be used as high-speed switches in computer circuits, or as electric oscillators, producing anything from 50 cycles per second, to 30,000 million cycles per second. A single tunnel diode can do the work of several conventional components. But if matter behaved in a completely predictable manner, the tunnel diode would not exist. It is a form of semi-conductor diode — two pieces of semi-conducting material joined together — which acts as a currentcontrolling barrier. In the conventional semi-conductor diode, the barrier is thick and impenetrable; it is a ‘wall’ caused by a difference in electric potential, and electrons cannot cross it. In a tunnel diode, the barrier is made OUTGOING WAVES INCOMING WAVES The odd behaviour of the tunnel diode shows that matter is governed by a form of statistics — wave mechanics. The current reaching the barrier can be represented by a wave. The probability of its getting through a substantial barrier is low. But the probability is higher for a thin barrier. so thin that electrons can tunnel their way through it. This shows that it is not zmpossible for electrons to cross a thick, ‘opaque’ wall — it is improbable. The behaviour of electrons in the their most probable means of crossing depends on the voltage across the barrier, and is controlled by the voltage. This kind of voltage control is especially useful in electronic switching circuits. In a typical tunnel diode, when the voltage difference across the barrier is a few hundredths of a volt, electrons are able to tunnel their way through. But by the time the voltage has been increased to about half a volt, it is sufficiently high for some electrons to jump over the top of the barrier. Between a few hundredths and a few tenths of a volt the barrier thickens — electrons cannot tunnel easily and still have not enough energy to jump. Increasing the voltage difference across a semi-conductor diode barrier has the effect of ‘lifting’ electrons on one side of the barrier, giving them more potential energy (another word for voltage is electrical potential). Negative Resistance The electrons are in fact lifted to a region where, although the voltage is rising, the current allowed through the barrier drops. An increase in voltage brings about a decrease in the current flowing. It is as if the diode had a negative resistance, a very odd state of affairs indeed. The graph showing the relationship between the current and the voltage difference has a pronounA TUNNEL DIODE ced dip in this region. The slope downwards is a negative slope, and corresponds to a negative electrical resistance. The slope upwards after the dip is an indication that electrons are beginning to pour over the top of the barrier. An increase in the voltage brings about an increase in the current. The resistance is positive again. Tunnel Diodes as Oscillators If a battery is connected to the tunnel diode so that the voltage across the barrier is midway between its tunnelling and jumping values, the diode starts to produce electrical oscillations. One Cycle of Oscillation in a Tunnel Diode. STAGE (1!) INCOMING ELECTRONS STAGE (2) ~, ELECTRONS DROP TO A LOWER ELECTRICAL POTENTIAL current is not entirely predictable — it is in fact governed by a kind of statistics, in which nothing is completely impossible, but some things are more probable than others. BATTERY BIASSES TUNNEL DIODE HERE ° A 5 roy. ft a . 4 In a tunnel diode, electrons trying 2 F Ni aie ee to get from one side of the barrier to 3 Y TO A MINIMUM the other have two possible routes. They can either tunnel through it, or summon up enough energy to jump over the top of it. They do not tunnel and jump at the same time, because VOLTAGE The current is at its lowest ebb. Jumping is The current puts its potential energy in impossible, tunnelling difficult. reserve, and starts to tunnel. 2076

During half of the cycle, electrons tunnel; during the other half, they jump, avoiding the negative resistance region, but obliged to cycle around it because the battery ‘biasses’ the barrier right in the middle of the negative resistance slope. At the start of the cycle, electrons tunnel, their numbers reaching peak value with a speed which depends on the amount of electrical ‘delay’, or inductance in the circuit. This flow of current has been produced by the battery, and has been given energy from the battery. Once the flow has been set in motion, it tends to keep going. The current ought to drop, because the diode is at the start of its negative resistance region. But it does not drop. Instead, electrons start to jump to maintain a constant flow, drawing in from reserves of energy. These are soon exhausted. The flow gradually drops down to its lowest point, right in the middle of the ‘dip’, where only a very few can jump and none can tunnel. At this stage, electrons might as well be tunnelling — this is certainly easier, for they can do it at a lower voltage. So electrons switch over almost instantaneously to tunnelling. It is at this part of the cycle that they put some energy in reserve. The tunnelling current then increases to a maximum, switches over to jumping, uses all the reserves, and so on. All the time the battery is making MAXIMUM FLOW THROUGH THE TUNNEL CURRENT RISES TO ITS MAXIMUM HERE - STILL TUNNELLING 4 ’ i CURRENT 4 _-”_ CURRENT HAS JUMPED <= @=— ACROSS FROM ITS MINIMUM The battery raises the voltage towards its biassing value. The current increases. STAGE (3) up for resistance losses in the circuit, by supplying energy. The diode switches from tunnelling to jumping practically instantaneously. The frequency of the cycles, or the number of complete cycles executed each second, depends only on the time taken for the current to rise to the maximum, and the fall to the minimum. This can happen very rapidly indeed, which is the reason why tunnel diodes can oscillate at frequencies of up to 30,000 million cycles per second. These frequencies are in the microwave region, and can be used for radar, communications, satellite control, and so on. But if electrical ‘delays’ are added, the frequency of oscillation can be made much lower. Tunnel diodes may be used in the future to convert low voltage direct current (from thermionic sources, one of the developing methods of power generation) to the low frequency 50 cycles per second alternating current used in the electricity mains. Present tunnel diodes can convert 60% of direct current energy, at only half a volt, to alternating current energy. Their maximum current output is only 6 amps. Higher efficiencies, and considerably higher power outputs, are needed before the tunnel diode becomes a practicable proposition as a high-power oscillator. MAXIMUM FLOW OVER THE BARRIEF STAGE (4) THIS LEVEL ENERGY RESERVES GONE To keep up the peak current flow, the energy reserves are used in jumping. THE BATTERY TRIES TO MAINTAIN THE VOLTAGE AT Materials for Tunnel Diodes All semi-conductor diodes are made from two pieces of semi-conductor material. One piece has had a kind of impurity added to it, which mobilizes ‘negative charges. In the other piece,. positive charges are freed. But, on the whole, both pieces remain electrically neutral. Charges on either side of the junction between the pieces influence each other electrically. Positive charges repel other positive charges, and attract negative charges. The result is that an excess of negative charge appears on one side of the junction, held there by an excess of positive charge on the other side. Together, the positive and negative charges act like an electrical barrier —rather like a battery which does nothing to stop the current flow in one direction across the junction, and hinders it in the opposite direction. The normal type of semi-conductor diode is made of germanium, with . about ten thousand million atoms of impurity added in each cubic centimetre to provide the mobile electrical charges. But this is a small amount compared with the impurity in tunnel diodes. They can also be based on the semi-conductor germanium, or on gallium arsenide. However, they contain between a hundred and a thousand million times more impurity per cubic centimetre. This has the effect of narrowing down the no-man’s land boundary region between the two ieces of semi-conductor. The barrier is very thin indeed, which is the reason why electrons can easily tunnel Electrons can tunnel through in both directions, so the tunnel diode does not act like a one-way barrier. Raa ced in CURRENT Once the reserves are used up, the current drops to a minimum-—where the cycle started. 2077

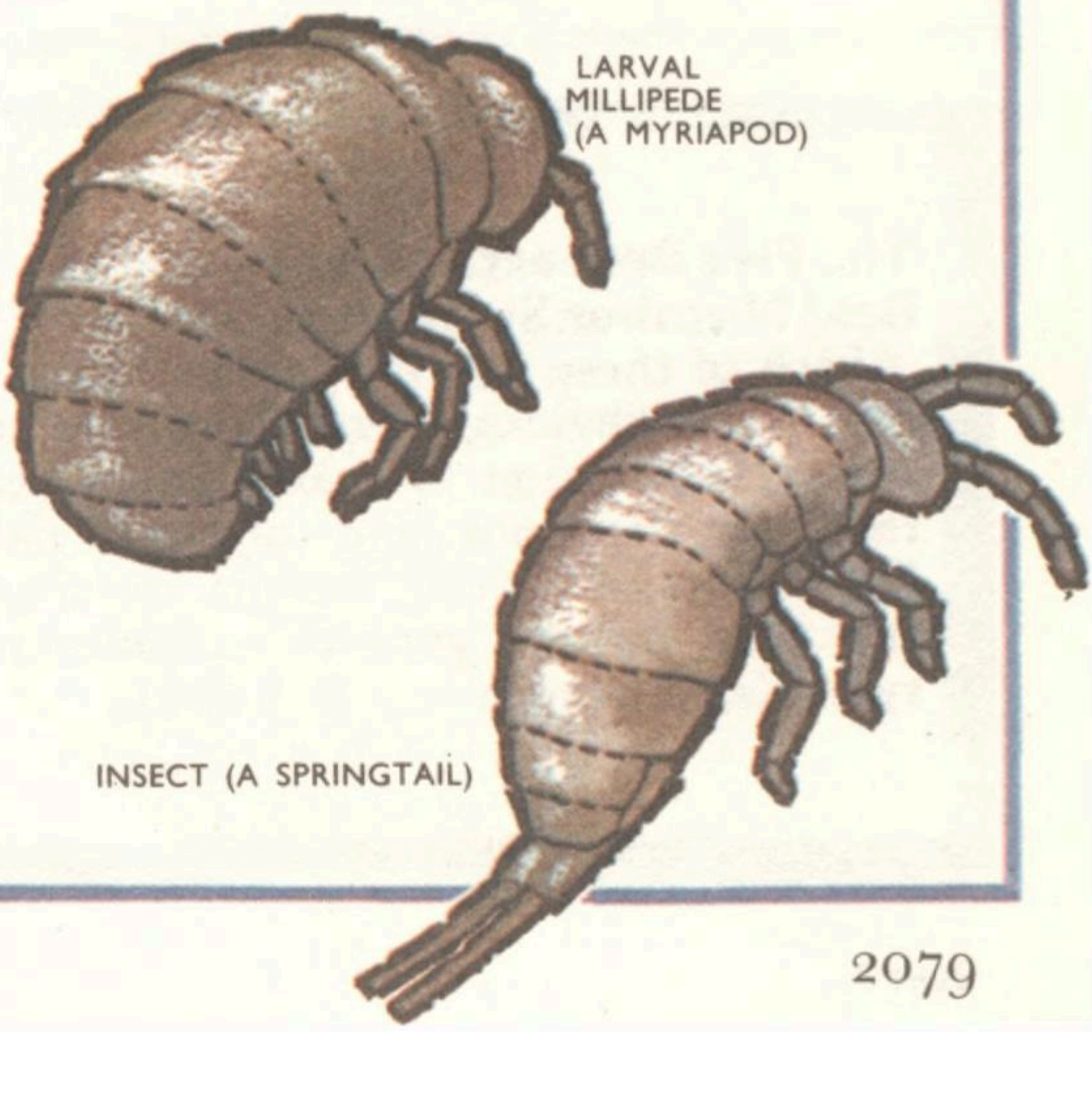
| BIOLOGY | PAEDOMORPHOSIS — New Animals from Old A "trange transformation astonished biologists in the year 1865. Certain types of 'giant newts' called Axolotls had been brought to the Zoological Gardens in Paris. Blunt nosed, dark in colour, equipped with three pairs of feathery external gills, these amphibians were known to spend the whole of their lives in freshwater. Unlike most amphibians they never came onto land. But the Paris specimens began to change. The external gills disappeared; so did the crest on the tail. Eyelids appeared and the colour of the skin altered. In place of an Axolotl, there appeared the familiar striped Tiger Salamander, a common amphibian throughout the eastern regions of the United States. Immediately it became clear that Axolotls were really a race of 'permanent larvae'. They lived to become sexually mature yet they never lost certain of their larval characteristics. The reason for this incomplete development was at first a mystery. Today the reason is thought to be due to lack of iodine in some lake waters . Other examples of Salamander larvae that do not usually complete their development, become recognized. Some such as Necturus the mud-puppy of the Eastern United States, are permanently larval in form. No matter what conditions surround them, they keep their larval appearance. The life of an individual animal 2078 usually commences as a single fertilized cell. By division the cell produces numerous other cells. These are arranged to form the tissues and organs of the new creature. The complete life history of the individual, from egg, through embryonic stages, and immature juvenile stages to the mature adult is called its ontogeny. Changes caused by recombination or alteration of genes can occur at any stage in the life history. Another factor that can change is the time and the Left, the adult tunicate or sea-squid ts rather sponge-like in appearance; but its larva — resembling a tadpole — has gill slits, a dorsal nerve chord and a notochord. By failing to develop into the adult, could not such a larva give rise to higher chordate animals such as the first fishes ? order in which various structures develop (heterochrony). Through the slowing down or speeding up of the processes of development, marked effects can be made on the final stages of the life history of an animal. Thus in the instance of the axolotl, TIGER SALAMANDER AXOLO Below, Axolotl, an am of the western U.S.A. Axolotl usually retains its larval appearance — for instance it keeps its external gills — even when sexually mature. It can however be induced to fully develop and changes dramatically into the Tiger

Salamander (above). certain larval characters are retained in the otherwise sexually mature animal. This retention is called Neoteny. An extreme case of neoteny — when the sexually mature animal is larval in every other respect is called Paedogenesis. The two processes Neoteny and Paedogenesis are commonly referred to as Paedomorphosis. The process of paedomorphosis has come to throw great light on evolution. In effect, new forms of animal can be produced from old forms. The appearance of a sexually mature descendant could differ markedly from the appearance of the sexually mature ancestor. Rather would the descendant resemble in appearance the young stage of its ancestor. From an evolutionary viewpoint there are great potential advantages. The larval structures, once they have become persistent, have an increased Left, feathers of the ostrich resemble the fledgling down feathers of other birds. They are probably neotenous features. So probably are the drooping ears of many domestic dogs, which resemble the ears of young wolf cubs. DOWN FEATHER PREDOMINANT IN FLEDGLINGS FEATHER OR ‘PLUME’ OF FLIGHTLESS OSTRICH YOUNG HARRIER WOLF CUB — EARS DROOPING OF DOG HAVING DROOPING EARS &



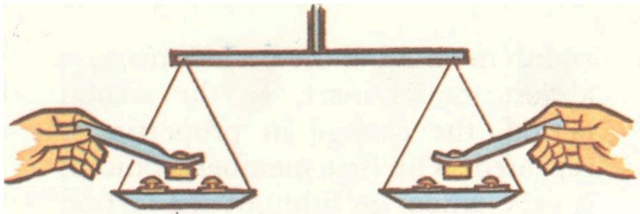
THE CHANGE CRESTED TAIL time in which they can further de- ' velop; new variations of the structure may appear. Because former adult characters no longer appear, genes of these lost characters are possibly available for new variations. By paedomorphosis and the sudden appearance of evolutionary novelties, completely new lineages could have branched off from older lines. Possible Origin for the Vertebrates Tunicates are tiny, barrel-shaped creatures. Usually they live anchored to the sea floor — adding to their already rather sponge-like appearance. The only prominent structural features to be seen from the outside are an opening at the top through which water is wafted and an opening at one side through which the water flows out. From the water passing through minute food particles are filtered. Certainly there seems little similarity between the adult tunicate and the backboned animals such as the fish and amphibians. But some tunicates possess a mobile larva. It is rather like a tadpole with a large head and an oscillating tail. The activity of the larva means that there is the possibility that new areas can be colonized by the tunicates. The internal structure of the tunicate larva is very interesting. Along the upper (dorsal) surface there is a hard but flexible bar called a notochord. Above it lies a hollow nerve cord, while in the head is a rudimentary brain. These are characters of all higher vertebrate creatures. Is the. relationship more than coincidental ? When the ‘tadpole’ finally settles down and develops into the adult tunicate these characters are lost. Should not, by paedomorphosis, the fixed way of life be lost altogether, the way would then be open for the ‘larva’ to give rise to the first fishlike vertebrates. Gaps in the Fossil Record Darwin’s theory of evolution explains how the more complicated organs could arise from simpler ones. A continuous series of organisms should theoretically have existed linking descendant stocks with their ancestors. The fossil record — the preserved remains of animals which have lived in the past — is the most obvious place to look for ‘missing links’. A few finds have been successful but generally creatures which link together whole divisions of the animal kingdom are rare. If paedomorphosis has played a part in evolution, as is likely, such ‘missing links’ could be accounted for. Paedomorphosis demands fossil gaps. By retarding developmental processes, quite abrupt modifications could take place. The millions of years required for evolution would not be necessary as there would be no gradual changes. Gaps between annelid worms and molluscs, are probably account ed for by paedomorphosis. Fossil evidence does in some instances support paedomorphosis. Young stages of trilobite preserved in ancient rock closely resemble in appearance the adult forms of later ~ trilobites. Similar evidence is found in remains of the fossil graptolites and ammonites. Paedomorphosis and Man The embryo of the anthropoid apes have the following characters — a relatively high brain weight in comparison with the rest of their body, scanty hair, and a flat face. Further the foramen magnum (the orifice with which the vertebral column articulates with the head) has a forward position underneath the skull. Consequently the head is balanced more upright on the spine than in the adult apes. Is there not something rather familiar about these characters? There is — they are all characters which distinguish Man. Theory of Recapitulation In 1866 Haeckel put forward his theory of Recapitulation or Biogentic Law. 'The life history of an individual' argued Haeckel 'recapitulates the whole of its ancestry.' For example, the embryo of Man passes from a single-celled state through stages resembling fishes, amphibians and reptiles. Haeckel's theory has been proved completely untrue. The embryo of Man never resembles the adult of any other creature that ever lived. But the embryo does bear a semblance to the embryos of other vertebrates; this denotes a related ancestry. There are occasionally instances which would seem to support Haeckel's theory. Sometimes adult structures can be pushed back into the young stages of a descendant. This is due to another process of heterochrony called Acceleration. Unlike paedomorphosis, acceleration is of no real evolutionary importance. Commonly, embryonic features persist into the adult animal with progressive development and change —a process of heterochrony called Deviation. When structures which function in the young stages are lost in the adult — for instance the tails of frogs — the process is one of Reduction. This close similarity between Man and a foetal ape is probably much more than coincidence. It seems extremely likely that in the remote past, the retention of early embryonic stages, with the consequent nonappearance of specialized adult features, led to the branching off of human stock from an _ anthropoidlike ancestor. The larvae of myriapods (centipedes, and millipedes) only have a few segments and three pairs of welldeveloped legs. At this stage they bear a strong resemblance to insects. Insects may well stem from such larval forms by paedomorphosis. LARVAL MILLIPEDE (A MYRIAPOD) INSECT (A SPRINGTAIL)

The larvae of myriapods (centipedes, and millipedes) only have a few segments and three pairs of well-developed legs. At this stage they bear a strong resemblance to insects. Insects may well stem from such larval forms by paedomorphosis.



MATHEMATICS The Negative Rules iT tia A) Subtraction | 1B 9 @ | | SOS e- oe @6@ eet) ——— 7 —6 This is not the same as The order of subtraction matters. Subtraction is not commutative. This is not the same as By The order of division matters. Divisio is not commutative. : 3 IS DIVIDED INTO 9 SEGMENTS The Five Properties of all Real Number Systems Each of these five properties looks, at first sight, obvious. Here the properties are set down first for natural numbers, then for fractions and ordered pairs (both ways of writing rational numbers) and then for the general numbers and symbols, written here as A, Band C. 2080 ball game without rules would result in complete disorder. And, in mathematics, there have to be set rules for playing with numbers. Some of these rules are obvious and self evident. We use them unconsciously because they always give the expected answer. But the rules are important — they define the game. Mathematical rules define the number system in operation. Real numbers — the natural numbers, the integers, the rationals and irrationals — use the same basic rule-book. They have five properties in common. Other kinds of quantity — imaginary numbers and vectors — follow a different set of rules, The five basic properties do not set out how the operations of addition and subtraction, multiplication and division are to be carried out. These vary considerably among the real numbers. The operations for rational numbers are more involved than the methods for natural numbers and _ integers, because the rational number is really a pair of numbers. Rules for addition on the denary (scale of tens) system are different from the rules of the binary (scale of twos) system. The five basic properties do not show how the game is to be played — they are rules of order and arrangement. The four basic operations are addition and subtraction, multiplication and division. Two or more of these may appear on the same line. Which operation is to have priority? Operations - carried out in different orders may give different results. Which is the right one? MATHEMATICAL RULES Adding and Subtracting Natural Numbers and Integers Addition and subtraction can be represented as movements along the number line, to right (addition) and to left (subtraction). This would work either for the scale of tens, or for the scale of twos. (4, 1) ae ADDITION AND SUBTRACTION IS STILL EQUIVALONG A NUMBER LINE (3, 3) Ordered Pairs The rationals, written as ordered pairs, are arranged along a number line. Addition is simplified when the pairs belong to the same set. Multiplying and Dividing Multiplying 8 x 12 is the same as adding 8 12 times. Dividing is the same as multiplying by one-over-the-number. It is the inverse of multiplication. Multiplication tables vary according to the number scale in operation — binary tables are different from denary tables. When numbers are written as ordered pairs, they can be multiplied and divided by multiplying numbers from one pair with numbers from another pair.

In equations, the golden rule is ‘fair play’ to both sides In other words, an operation carried out to one side (perhaps to simplify the equation) must must be preserved. To avoid confusion, the rule-book lays down that multiplication and division have priority over addition and subtraction, and, all other things being equal, should come first. If, however, it is essential that addition come before multiplication, the addi- « tion is enclosed in brackets. Brackets are a method of showing when the basic rule-book properties are amended. The operation inside a pair of brackets must be carried out before any others. This priority rule is included in the distributive property of number systems. Is 4 + 3 the same as 3 + 4? Or is 4 —3 the same as 3 —4? In the first example, it is the same, and in the [x |+[@e! [ae] second it is not. When numbers are to 283 e\i- +/8 is EQuWALeN teeel= be added, their order does not matter, and can be changed around. But num- Stage (1) Dividing both sides by 2. _ ~ pC i=[x | Soe i Oo oO! r bers being subtracted cannot be Stage (2) Subtracting 2 from both sides. —— Poretoicnt changed around. Addition is said to be commutative ; subtraction is not. In the same way, multiplication is also commutative, division is not. 3 x 4 equals 4 “a : : i : 7 ae ae fy 5 Obeying the ‘Fair Play’ Rules —- Multiplying and Adding + 7?Inthe first, the operation 5 + 7 is carried out first; in the second 2 + 5 is (1) |@ @ carried out first. Both give the result lo] a xg — ip” | sien fe] dab g + [x] = — : os) o also be done to the other side. The balance (signified by the ‘equals’ sign Obeying the ‘Fair 33] Rules — Dre and Subtracting 14. So the grouping together of num- 4 2 1S EQUIVALENT bers for addition does not matter, and nor does it matter for multiplication. Stage (1) Multiplying both sides by 4. Again, subtraction and division are (2), (3) 29 excluded. Addition and multiplica- Stage (2) Adding 4 to both sides. @ tion are said to be associative, while gases 2 ° = [x] subtraction and division are not. Stage (3) Subtracting x from both sides. © EQUIVALENT ae Two commutative properties, two associative properties (both for addition and multiplication) and one distributive property together make up the five properties of real number systems. Multiplication is Distributive over Addition —or unless brackets indicate otherwise, multiplication comes first. The priority rule is: brackets, multiplication, division, addition, subtraction. 3+4x5=3+20=23 But J mei ea =7x5=— “4 2081



Adding numbers to both sides of an equation is like adding the same weights to both sides of a balance. Similarly, the same weights can be subtracted from both sides.

Destroying the balance by adding (or subtracting) more to one side than the other. Balanced equations produce **equivalent statements,** unbalanced equations do not.

Addition is Associative

$$3 + (4 + 5) = (3 + 4) + 5$$

$$\frac{1}{3} + \left(\frac{1}{4} + \frac{1}{5}\right) = \left(\frac{1}{3} + \frac{1}{4}\right) + \frac{1}{5}$$

$$(1,3) + \{(1,4) + (1,5)\} = \{(1,3) + (1,4)\} + (1,5)$$

$$A + (B + C) = (A + B) + C$$

Multiplication is Associative

$$3 \times (4 \times 5) = (3 \times 4) \times 5$$

$$\frac{1}{3} \times \left(\frac{1}{4} \times \frac{1}{5}\right) = \left(\frac{1}{3} \times \frac{1}{4}\right) \times \frac{1}{5}$$

$$(1,3) \times \{(1,4) \times (1,5)\} = \{(1,3) \times (1,4)\} \times (1,5)$$

$$A \times (B \times C) = (A \times B) \times C$$

INORGANIC CHEMISTRY The TRANSITION ELEMENTS

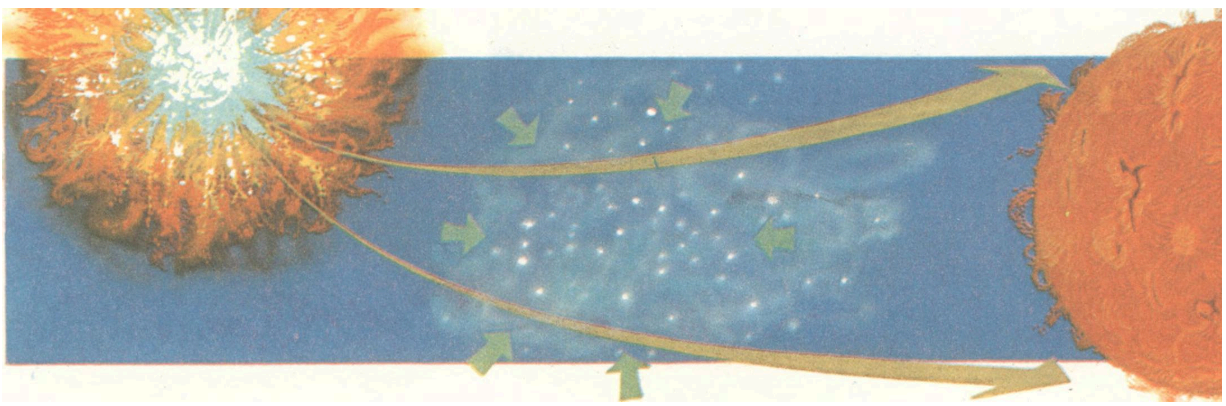
ERE are now 104 different elements known to Man, and the whole of the Earth is thought to be composed of these elements either in the pure form or in various states of combination. The properties of the elements vary enormously — ranging from highly reactive gases such as fluorine and chlorine and highly reactive metallic solids such as sodium and potassium to the inert gases and the heavy unreactive metals such as silver and gold. Since the earliest days of chemical science, attempts have been made to put some sense and order into the classification of the elements — to explain the wide differences in the properties of some elements and the similarities of others. This was eventually done, with outstanding success, by Dmitri Mendeleeff when he devized his Pertodic Classification. He found that if the elements were arranged in order of increasing atomic weight, they fell into natural periods. The two lightest elements, hydrogen and helium, form a short period of their own. These are followed by the period of eight elements starting with lithium and ending with neon. The next period of eight elements starts with sodium and ends with argon. When the elements were arranged in this way, they formed natural families. In each period there was a gradual change in properties from one element to its neighbour. In the first period, the first member, lithium, loses an electron very easily — in other words, it is strongly electropositive. Proceeding through the period, the elements become less electropositive until the seventh member is the strongly electronegative fluorine. The C WA phn pram = ed ; ‘ : ! 4 f 2082 eighth member of the period, neon, is almost totally inert. In the second period, the change in properties is repeated. The first member, sodium, is very similar to lithium; the second member, magnesium, is very similar to the second member of the first period, beryllium. This repetition of properties carries on through the second period, and the periods that follow it. Although Mendeleeff did not realize it at the time, he had arranged the elements according to the numbers of electrons each element has in the outer shell of its atom. Thus, lithium and sodium each have one electron in their outer shells and are very similar chemically to the first member of the third period, potassium, which also has only one electron in its outer shell. Moving along each of the early periods there is a build-up in the number of electrons in the outer shells of the atoms. When the last, inert element (neon, argon etc.) is reached there are eight electrons in the outer shell. All the chemical properties (and many of the physical properties) of the elements are decided by the arrangements of the electrons in the shells. The outer shell is particularly important because it contains the valence electrons. These are the electrons that can be given to other atoms to form electrovalent bonds, or shared with other atoms in covalent bonds. The number of electrons in the outer shell is the main factor that decides what sorts of chemical reactions the element undergoes and how it becomes bonded to other atoms. So, by arranging the elements in the periodic classification, Mendeleeff automatically sorted the elements into groups with equal numbers of ‘outer shell’ electrons. This also meant that the elements were classified according to chemical properties. The ‘octave’ classification with the number of electrons increasing until the ‘magic number’ of eight electrons is achieved works very well for the elements of small atomic weight, but it falls down in the fourth period of elements. In the earlier periods there is the expected build up of electrons in the outer shell, from one to eight, but in the fourth period this pattern is interrupted. The first element, potassium, has a single outer shell electron. The second element, calcium, has two electrons. The third element, scandium, would be expected to have three electrons but, in fact, has only two. Most of the next nine elements in the third period have only two electrons in the outer N shell. The atoms of the elements have extra electrons but instead of going to the outer shell they are going to the next inner M shell. So scandium has only two electrons in the N shell, but has nine in its M shell. Its neighbour titanium has ten electrons in its M shell, and the build-up of electrons in the M shell continues until zinc is reached. This has eighteen electrons in its inner M shell. After zinc, the original pattern, with the build-up of the outer N shell of electrons is continued. The last elements of the period all have eighteen electrons in the M shell and they progress, one electron at a time until eight electrons are acquired by the outer N shell. All the elements from scandium to zinc are called transition elements. In the fifth and sixth periods of the periodic table the build-up of the outer shell is similarly interrupted. Further families of transition elements are formed by the build-up of electrons in inner shells in these periods. The outer shells contain only two electrons and the elements have many properties in common. Properties of the Transition Elements Many of the most common and important metals are transition ele

@ ©) xz Oe x © x IN BUILD-UP OF ELECTRON SHELLS IN TRANSITIONAL ELEMENTS ELECTRONS ARE ADDED TO INNER SHELL ments, including iron, copper, zinc, silver and gold. All of the transition elements are dense shiny metals — good conductors of heat and electricity. Included in the transition metals are the strongly magnetic (ferromagnetic) metals iron, cobalt and nickel. Most of the other transition metals are paramagnetic — weak versions of the ferromagnetic materials. These magnetic properties are due, once again, to the arrangements of electrons in the shells. In ferromagnetic materials, for example, there are unpaired electrons in the shells — for one electron spinning or a 2 yi 2 2 2 a ys EXTRA ELECTRONS PASS OUTER (L) SHELL ay @ oa ; LETTERS K, L, M, N, REFER TO ELECTRON SHELLS, AND NUMBERS REFER TO NUMBERS OF ELECTRONS IN THOSE SHELLS rc xz K L L LM Kt 8 28 2 8 an MN KLMN KLMN K LMN 2 oy Bn in one direction there is not another spinning in the opposite direction to neutralize the magnetic effects produced by the first. The chemical properties of these metals are very largely affected by the tendency of the electrons in the inner shell to take part in chemical bondings. Apparently, there is an 'urge' on the part of these electrons to behave in a similar way to the valence electrons in the outer shell. They are similar in behaviour to the electrons in the outer shell because they, too, are not occupying a complete shell of electrons. For example, an atom of the transitional element iron may lose two electrons from its outer shell to form the ion The first part of the periodic table, showing the intervention of the first family of transition elements into the fourth period. The electrons are gained by an inner shell, and not the outer shell, as expected. Fe+*, or, in addition, lose a third electron from the next shell to form the ion Fe++. The main factor in determining the chemical properties of these elements is, however, the electrons in the outer orbit, and in passing along the IN 'BUILD-UP' OF ELECTRON SHELLS ELECTRONS ARE NORMALLY ADDED TO OUTER SHELL period from one element to the next the properties of neighbouring transitional elements change very little. The transitional elements all form positive ions, usually highly coloured both in the solid and in solution. Some of them also form complex ions. The metallic ion takes up molecules of water or ammonia, or even other ions and chemical groups (e.g. the cyanide ion CN') to form these coordination compounds. A well known example of this is the cuprammonium ion (Cu4NH3;)** formed when ammonia solution is added to a solution of a copper salt. The deep-blue solution that results is used to confirm the presence of copper ions in qualitative analysis. 2083

| HEAT PHYSICS | The Porous Plug Experiment A gas molecule can never be seen, even under the most powerful microscope, but the scientist has developed a fairly clear picture of what it is like. He knows, for example, how the atoms that make up a molecule are bound together, and he also knows quite a lot about the internal structure of the individual atoms themselves. THERMAL i The gas is kept at a constant high pressure on one side of the porous plug and expands to a constant lower pressure on the other side. The molecules have to overcome forces of attraction and this produces cooling in the gas. But, of even greater practical importance is the behaviour of large numbers of molecules or atoms when they come together to form a gas, liquid, or solid. The physical properties of a substance are measured in the laboratory and from these measurements, physical laws are derived. These are macroscopic laws. — they describe the properties of vast numbers of molecules or atoms in bulk. These laws are used by the scientist and engineer to help predict how the substance will behave under different conditions. They can also help confirm the scientist's theories of what processes take place on the microscopic scale. For example, the macroscopic properties of gases, the effects of changing temperature and pressure, can be studied in the laboratory and are described by the gas laws. These laws help the scientist confirm the theoretical, microscopic 'model' — the kinetic theory of gases. 2084 According to the kinetic theory, the molecules in a gas are in constant motion, and the pressure exerted by the gas on the walls of the container is due to constant bombardment by atoms and molecules. From this theory the ordinary properties of the gases described by the gas laws can be predicted. But there is one factor that is not taken into account in the simple kinetic theory. This is the forces of attraction existing between the gas molecules. This factor leads to inaccuracies in the gas laws, but by conducting experiments, the effects of the attraction between molecules in a gas may be studied. This was first done by James Joule and Lord Kelvin (William Thomson) in a famous demonstration which is now sometimes known as the porous plug experiment. A porous plug of absorbent cotton or silk was placed in a tube and a constant pressure difference between the two sides of the plug was established so that gas was forced through. On passing through the plug the gas expanded. The whole of the apparatus was very carefully lagged to ensure that no heat could enter or leave the tube. Such an expansion, with

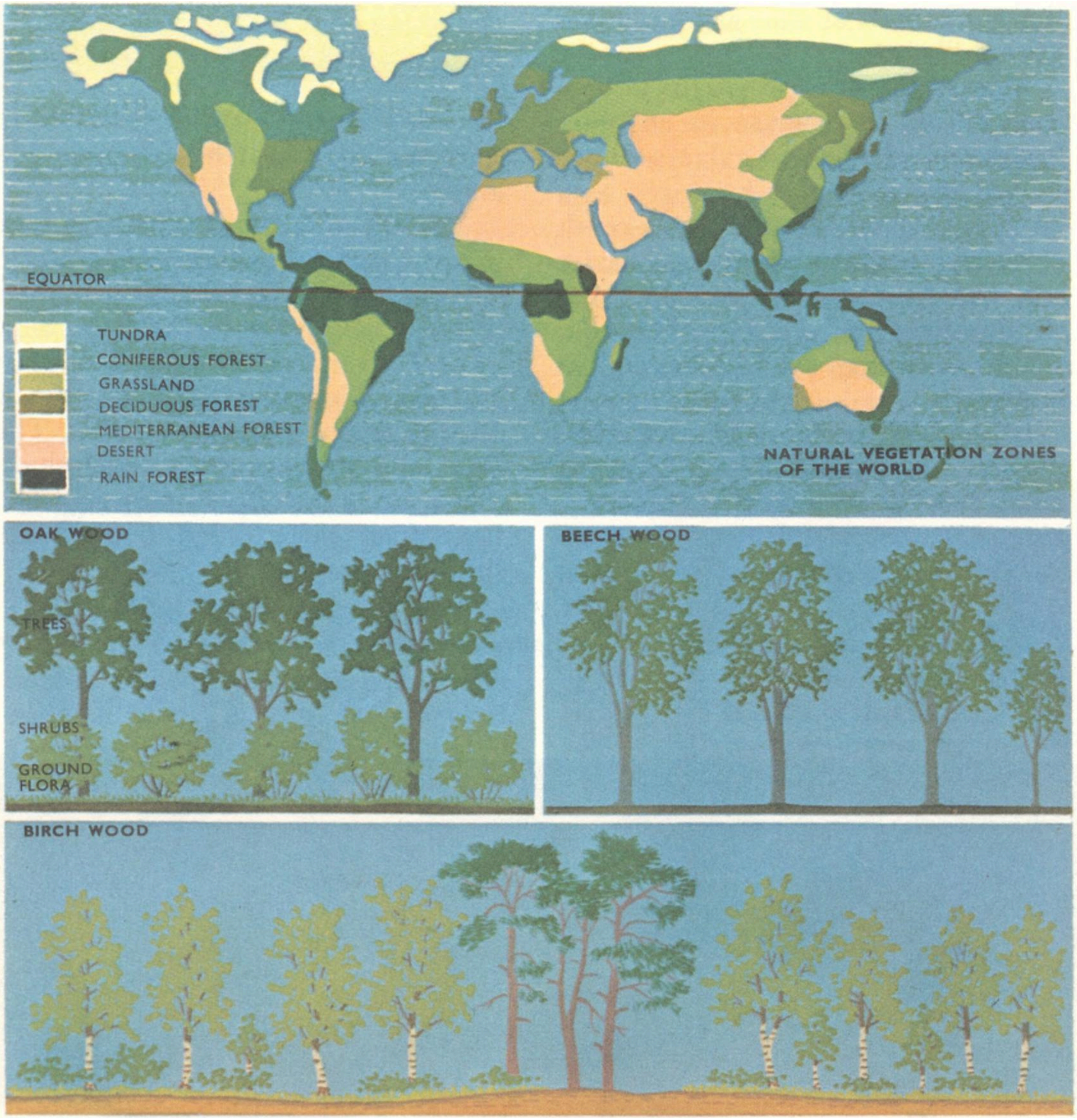
no heat transfer, is called an adiabatic expansion. It was found, for most gases, that as a result of the expansion, the temperaAIR IN ° TO MAKE UP FOR REMOVED LIQUEFIED AIR ture of the gas was reduced. The explanation of the cooling is as follows. Energy is given to the gas molecules when they are compressed on the high pressure side of the plug, but they lose energy when the gas expands on the other side. The compression-expansion process can lead either to heating or cooling depending whether there is a net loss or gain in energy. In addition, energy is lost when the molecules overcome the inter-molecular forces of attraction. The drop in temperature (the Joule-Thomson effect) occurs there when there is a net loss of energy as a result of both processes. After the original experiments, other workers repeated the experiments using more efficient porous plugs, made of earthenware and _ other materials. They confirmed Joule and Thomson's results, but also found that not all gases produced a cooling effect at all temperatures. Helium and hydrogen, for example, heated up when they expanded through the plug at ordinary temperatures, but cooled on expansion if they were first cooled below the inversion temperature. This is because there is a net gain in energy at the higher temperatures (but not at lower temperatures) — the gain in energy on expanding through the plug exceeds the loss caused by expansion against intermolecular forces. _ Inthe Linde process air is compressed and the heat of compression removed before the _ compressed air is passed through a nozzle. Joule-Thomson expansion takes place, cooling the air, which passes up the surrounding tube and back to the compressor to repeat the process. Eventually the air is liquefied and collected.

ASTRONOMY Supernovae and the Origin of Iron HE chemical composition of stars, such as the Sun, can be found by examining their light with a spectroscope. When the Sun's light is split into its spectrum with a simple prism, it appears to be an unbroken band of all the colours of the rainbow. Closer examination with a spectroscope reveals that it is crossed by dark lines, Fraunhofer lines, marking the wavelengths of light which have been absorbed by atoms in the outer, cooler layers of the Sun's chromosphere. Certain elements absorb certain wavelengths: so it is possible to identify the elements present in the light-emitting chromosphere. The relative intensity of the light and dark lines shows the relative abundance of the elements in the chromosphere. No-one will ever be able to probe deeper than this into the Sun; calculations show that if the chemical composition is uniform Two stable elements, helium and tron, play an important part in supernovae. HELIUM ATOMS IRON ATOMS throughout the layers of the Sun, then its mass is 73% hydrogen and 25% helium. Hydrogen and helium are the two lightest elements. The rest of the elements between them make up the remaining 2°% of the Sun's mass. Only one atom in 50,000 is of the element iron. But the presence of even this minute amount is difficult to explain. It is thought that iron can be formed only during a supernova (a stellar explosion). Supernovae in our own galaxy are brilliant events, easily observable. Records show that there have been two supernovae in the past 1,000 years. At a rate of one per 500 years, each supernova would have to create a massive amount of iron in the galaxy. A rough estimate shows that each would have to produce five times the Sun's mass of iron alone. This is very unlikely. Most of the iron is thought to be a legacy from supernovae in an older generation of stars and galaxies. They started out with only hydrogen as a building block. Gradually, as a result of thermo-nuclear reactions, heavier and heavier elements were built up within the stars, using protons and electrons from the hydrogen as material. In some of the more massive stars, the reactions got out of hand, and too much heat was released too quickly. Iron, a very stable element, could be formed. Then, suddenly, conditions favoured the break-up of iron nuclei into the next most stable element, helium. A vast amount of energy is required to break up iron. The star undergoes a collapse to provide the energy from its stockpile of gravitational potential energy. It becomes violently unstable, and explodes, as helium is re-converted to iron. Up to 90% of the star's mass is flung out into space, to form part of the material for the next generation of stars. Our own Sun is a member of this next generation of younger stars. So, in building up its present stock of heavier elements, it did not have to start out from scratch with hydrogen alone. There is no means of checking how many _ iron-producing supernovae occurred in the thousands of millions of years before the birth of the Sun. (The Universe as a whole is at least 10,000 million years old: the Sun is around 5,000 million years old). The remnants of these stellar explosions, dense, insignificant white dwarf stars, have long since faded beyond the limits of observation. Tron produced during a supernova (left) is ejected into space, and finds its way into younger stars (right). 2085



| ECOLOGY| aranatata . we Wee waa Bae & * AN oakwood in springtime: close to — the ground clumps of moss, white anemones, lesser celandines, winter aconites, and thick green spreads of dog's mercury; above, an undergrowth of stouter woody shrubs such as hazel, hawthorn and sloe; higher still branches of the oak trees themselves, supported on tough rigid trunks. All these plants are growing on a similar sort of soil; all are subjected to the same sort of climate. Together they make up a recognizable community of plants. The term plant community is of great use in plant ecology. It can be applied to any collection of plants which make up a distinct type of vegetation — from woodlands to the scant plant growth on the flanks of a sand-dune. For greater precision of meaning, according to the scale of the plant community, a number of other terms can 2086 be used. A plant formation refers to a community in very broad terms. The great belts of vegetation found throughout the world—the rain forests, desert vegetation, deciduous forests, coniferous forests — are examples. Studied in greater detail each plant formation breaks down into a number of subsidiary categories. Oakwoods, for example, are a smaller community falling within the larger community of the deciduous forest plant formation. For such a community, dominated by a single species (i.e. the oak tree) the term consociation is used. If there are two or more species of equal importance, then the community is called an association (e.g. mixed oak-ash woods). Finally, very small but distinct communities may occur inside associations or consociations. Within an oakwood for instance. there may be a local predominance of we 4 ht © Ae ge aE Trem" a | | s | : % NETTLES of p ANTS . & ra Bg 4 ae) 4 A , ash trees. This lowest category is referred to as a_ plant society. Distribution of Communities Why are rain forests found only in tropical regions? What causes the beech wood consociation to dominate the chalk and limestone hills of Southern England while oak woods cover the lowlying valleys? The occurrence and distribution of any plant community rests with three sets of factors. There is the climatic factor — including the influence of sunlight, temperature, wind, rainfall and humidity; there is the soil or edaphic factor — the composition and properties of the soil supporting the plant community. And there is the biotic factor — primarily the influence of the animal population on the plant community. The climatic factor is without any doubt the most important of the three. It is the world-wide variation in climate that gives

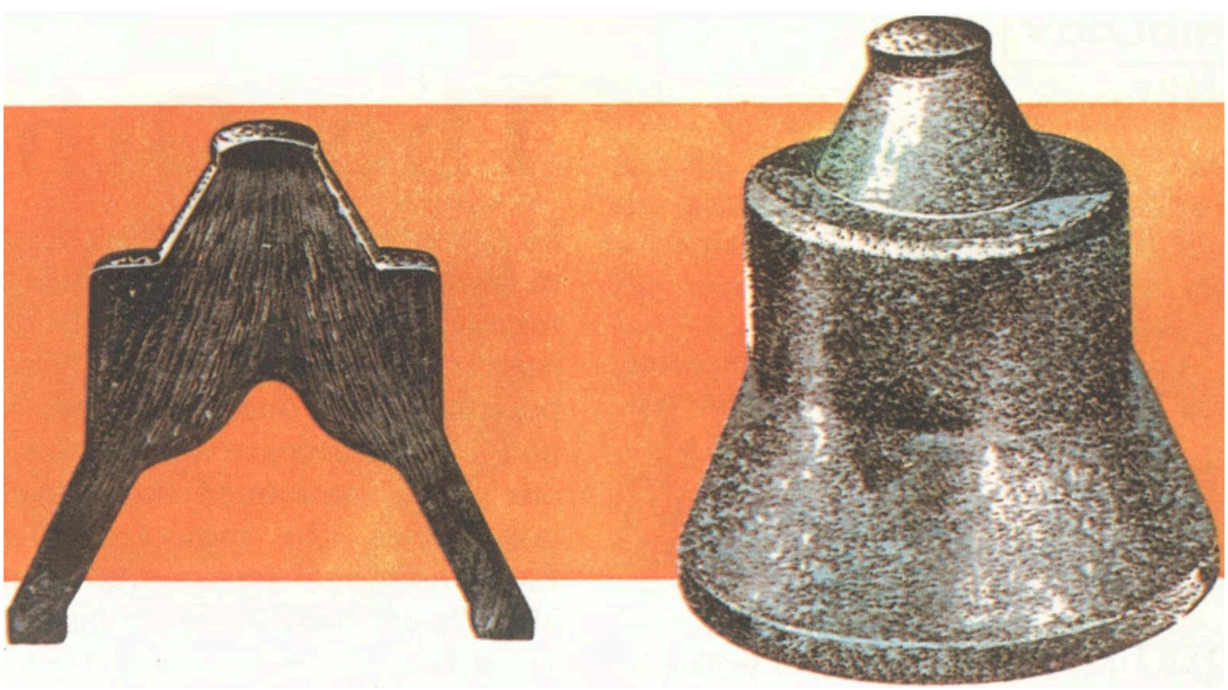
the characteristic vegetation belts or plant formation. No matter what the soil is, a rain forest would never grow in Western Europe. Rain forests need moisture, warmth and strong sunlight throughout the year and only in the tropical regions are those needs satisfied. Western Europe, however, with its warm moist summers and coldish winters is the ideal for deciduous forest. The soil or edaphic factor has a strong secondary influence on plant communities. The types of associations and consociations within a plant formation are largely determined by soils. Within the deciduous forest category for example, oakwoods usually monopolize the lowlying woodlands. The reason is that oaks are suited to the moist heavy clays generally forming low areas. Beech woods Above, plant formations of the world. Climate is the controlling factor. Middle /, oak-wood conso¢ n showing strafied structure of con (y. Right, beech onsociation — the) shade cast by ese trees prevent lower levels of plants rom forming. Below, a plant\society — localized occurrence of pines within a predominant birch forest.


EQUATOR
TUNDRA
CONIFEROUS FOREST
GRASSLAND
DECIDUOUS FOREST
MEDITERRANEAN FOREST
DESERT
RAIN FOREST
NATURAL VEGETATION ZONES OF THE WORLD
OAK WOOD
TREES
SHRUBS
GROUND FLORA
BEECH WOOD
BIRCH WOOD

NY a Nei = Beech trees cast a very deep shade. XY) BAN a) (t SR Ee, et Te RS oe se : = Consequently, in such a wood, the oN at ea Paty Sa vegetation nearer the ground is scanty . ~ or absent altogether. Oaks and ashes casting less shade enable far more abundant and varied undergrowths and ground floras to flourish. Apart from the differences in light intensity, slight variations can also be detected in the air's carbon dioxide, oxygen and water content at each level of a plant community. Thus superimposed on the broad climate of the community as a whole, each plant layer has its own peculiar micro-climate. SUNTAN Duneland plant communities are tolerant of poor soil with low mineral content. Here marram grass, sea spurge and sea holly begin colonizing shifting sand. More mature 'fixed' dunes are seen in the background, where plant succession has proceeded. favour light, shallow, limey soils and hence grow scattered over downs and limestone hills. Shallow sandy, welldrained soils favour the growth of birch trees and pines. Oakwoods may be found here as well, but this oak tree—the Durmast oak or Quercus petraea—is a different species to the Pedunculate oak (Quercus robur) of the lowlying clay soils. The biotic factor theoretically means the action of all organisms on the plant community, including the influence of the plants upon one another. But usually the phrase is taken to mean the effect of the animal population alone. It includes soil-dwellers such as earthworms, bacteria and viruses, the pollinating insects, destructive grubs, the browsing and grazing animals such as deer and rabbits, the seed-dispersing birds and of greatest importance of all — Man himself. Man's impact on plant communities with his axe, plough and herds of grazing animals is immense. For instance three thousand years ago Britain was completely wooded — that is all but high mountains, bogs and swamps. Centuries of agriculture have pushed the forests back to today's scanty woodlands. In their place are highly woodland. artificial plant communities — fields of crops and pastures carefully tended by Man to prevent unwanted, useless plants (weeds) from encroaching. Structure of Communities For sheer size, the tree dominates the plant kingdom. Its tough woody stem enables leaves and flowers to open scores of feet above the ground. But trees dominate in one other way. Their leaves, supported high in the air receive the maximum amount of sunlight available. Considerably less light penetrates the trees' foliage to reach the smaller shrubs, and _ less still reaches the herbaceous plants

growing near to the ground and the mosses actually in contact with the earth. This layering effect of stratification is a characteristic of all plant communities. In woodlands it reaches its greatest development with four ‘stories’ — trees, shrubs, herbs and mosses. Outside the woods, the layering may not be so complete yet nevertheless is still present. Grasslands, for instance, consist of an upper layer of tall grasses and herbs while beneath are rosette plants and short grasses. The tree as a dominant plant in a woodland community strongly influences the lower levels of growth. Plant communities evolve. From simple beginnings they may pass through a series of changes (the plant succession) until they reach an advanced woodland community. Imagine an area of land cleared of its top-soil by a bulldozer. The first plants to recolonize the area are hardy pioneers requiring little in the way of nourishment from the soil. They include the lowly algae, the mosses and the lichens. With the successive death of a few generations of these pioneer plants the soil has become meber in humus. In come new species of plant — mostly annuals with efficient dispersal mechanisms — that now can flourish in the better soil. They coun the pioneers, but in turn may largely be replaced by later perennial plants. Shrubs and small trees appear. With enough time, larger trees may become established and begin to dominate the community. The final result is a stable community in which no further great changes take place. The end-point is reached and this community is called the climax community. But the climax community is only stable so long as the surrounding conditions remain constant. Changes in the climate, in the soil, or changes caused by Man or other animals can upset the stability. The dominant species may decline and disappear and new species take over. Such an event has taken place in Western European woodlands since the end of the Ice Age. Obliteration of a pool. Swamp plants are succeeded by shrubs and small trees. The plant succession will end in a climatic community of

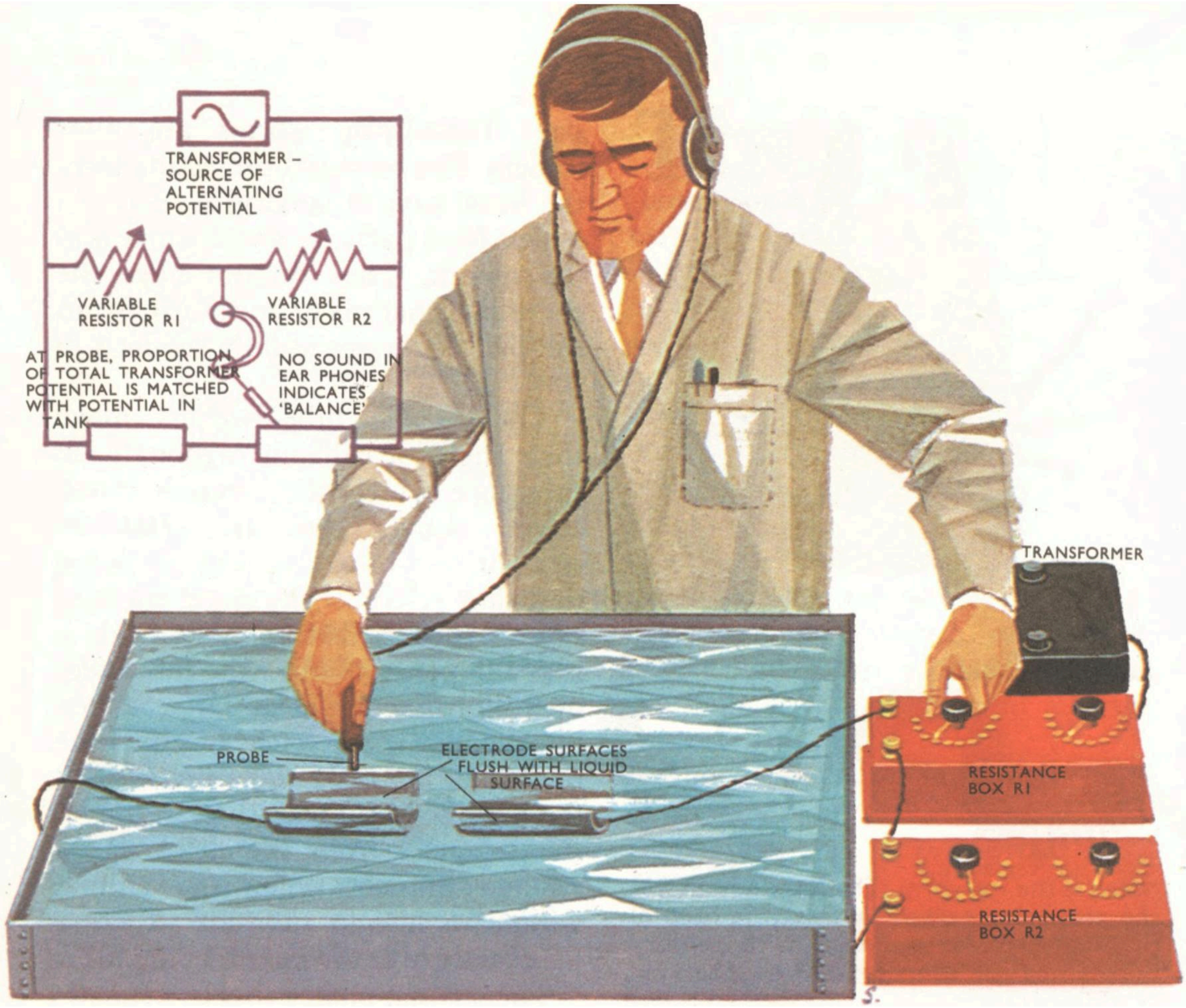
TECHNOLOGY 2088 HE forging of metals is an oldestablished art — practised over the centuries ever since man_ first worked metals into tools. The basic process — heating the metal until it becomes relatively soft and then forcing it into shape with a hammer or die —is the same as has always been used by the village blacksmith. In fact, hand forging as carried out in the smithy is still used in different sorts of engineering work, although enormous machines have been developed for forging large metal components. In these machines, steam or hydraulic pressure systems are used to forge the work-piece. In forging, the metal is deformed at a temperature higher than the recrystalization temperature and an important consequence of this is that no strain hardening (increase in resistance to deformation) results in the metal. Any non-metallic impurity in the metal is spread out to produce a fibre-like structure. The work-piece then develops a great toughness and ductability (the ability to be ‘drawn out’) in the direction of the ‘fibrelines’. There are a number of different categories of forging. Hand forging has given way, for larger components, at any rate, to hammer forging. The anvil holds the lower die, in which the gfal is forged to the required shape. is, might be flat, V-shaped or round. The molten work-piece is placed on the anvil and the upper die placed on the work-piece where a series of blows is delivered by a heavy steam hammer. There is a limit to the pressure that can be applied by hammer, and the largest forgings are made using hydraulic presses. Enormous pressures are developed in these but the work-piece undergoes a squeezing action rather than the impulsive ‘hammer-blow’ action of Tools used in forging. (Top) Traditional blacksmiths anvil. (Centre) A power hammer used in the rough forging of heavy components. (Bottom) A large hydraulic machine that shapes heavy components, using a succession of high pressure ‘squeezes? . the hammer. Hydraulic forging machines are the only ones capable of being used in the manufacture of the heaviest metal components. Forces of up to 50,000 tons may be developed in these. In some hydraulic equipment, in addition to the vertical ram, horizontal rams can be deployed to produce simultaneous pressings in several directions at once. Another important class of forging work is drop-forging. The lower die forms an integral part of the forging machine. The impact blow is delivered by a weight-ram that holds the upper die. There are two categories of drop forging machine. In the board drophammer type the weight is attached to wooden boards which are raised between rollers. The weight is lifted and held up by friction in the rollers and then allowed to drop and accelerate under gravity to deliver the impact blow. In the steam-hammer, the ram is actuated by means of a steam cylinder and falls under the combined influence of the steam pressure and gravity. Very high forging forces (up to 50,000 lb. wt.) can be developed in these machines. Drop forging is particularly useful when dealing with metals that require rapid treatment after heating. Some types of alloy can be heated through a small temperature range only, without High pressure gas fills the system and 1s further compressed by pistons that are forced downwards when oil is introduced. OIE INERODUCED losing some of their more desirable properties. Rapid treatment in the drop-forging machine enables the hot metal to be transferred directly from the furnace and immediately receive the forging treatment. Another advantage of the drop-forging method is that the work-piece needs very little ‘cleaning up’ to achieve its final form. High Energy Forging One of the most spectacular developments in recent years has been the introduction of impact forging, sometimes known as high energy rate forging. To obtain the greatest efficiency in the forging process the workpiece must be subjected to rapidly delivered impact blows. The maximum energy must be delivered to the work-piece in the shortest possible time. In the conventional steam hammer and drop-forging methods considerable amounts of energy are delivered to the work-piece in a short time, but in impact forging this is improved upon. The energy delivered by a hammer head depends on the mass of the head and on the velocity which it attains prior to impact. But the work-piece can be made to travel towards the hammer and the hammer travel towards the work-piece at the same time, so that the effective (relative) velocity of the hammer and workpiece is increased, and the amount of energy delivered to the work-piece This hub, formed from a solid cylindrical steel billet, is a typical high-energy forged product. Note in the cross-sectional view, the grain pattern produced in the metal. much increased. This is the principle of the wmpacter forging machine. In some impacter machines the workpiece is held in a stationary position and two rams are ejected so that they come together in the plane of the work-piece. In other systems, the work-piece is placed on one of the two moving platens and once again the platens are thrown together producing a large impacting force when they meet. The platens are energized by the application of high gas pressures in these machines. In one type of impacter machine, the platens are connected to each other by trombone-like rods which are expanded when the platens are parted and contracted when they Left. Trigger valve is closed, so gas in pressure chamber ts sealed off. But oil pressure is relieved and gas in trigger passages forces floating pistons upwards. Trigger valve automatically opens so gas can force up main piston and open up main seals. High pressure gas rushes in (right) and forces up main pistons — platens come together at high speed. Ram mee STR “ROS . | Cia i — : PISTONS RISE a BAS OIL [——_ SPR R 7 ore eee q aed : ue j : WaA YA are brought together. A charge-of gas (nitrogen) is introduced into a high pressure chamber and the two actuating cylinders at a pressure of 600 pounds per square inch (p.s.i.). This gas is further compressed by the introduction of oil into the cylinders, forcing a pair of pistons downwards under the oil pressure. The movement of the piston compresses the gas further and a gas pressure of 1,500 p.s.i. is thus developed in the high pressure chamber. A trigger valve mechanism is then operated and allows the high gas pressure under the floating piston to lift them to the previous, 'stand-by' situation, but at the same time seals off the high pressure chamber. The driving pistons are then situated between the higher pressure of the pressure chamber side, and a lower pressure on the cylinder side. This makes the pistons move with great rapidity and causes the two platens to be thrown together until they meet, at a relative velocity of 45 miles per hour. One of the great advantages of this method is that very little energy is wasted in 'pounding' the machine frame as in forging methods. All the energy developed is delivered to the work-piece itself. This makes for a clean efficient forging operation with very little surplus metal to be removed afterwards, and largely eliminates vibration and noise. It incidentally leads ta great economies in installation, because the special concrete mounting normally employed for large forging machines is not necessary. 2089

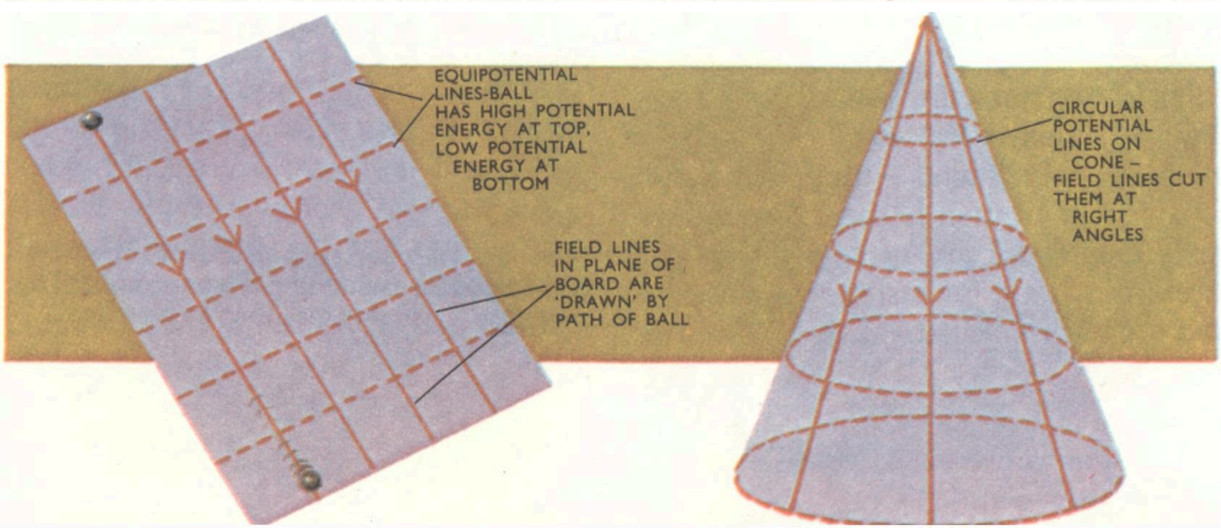


7 Sabella — a __ tube-dwelling worm. Beautiful feather-like tentacles, well equipped with cilia, emerge from .the opening. The tube — made of mud — may be a foot long, keeping the tentacles well clear of the sea floor. . BEST known of the annelid worms is . the common earthworm. It disjj plays the characteristic annelid ! features. Its body is divided into a | large number of segments basically all » identical in structure. For instance, 4 each segment has its own pair of excretory organs — the nephridia — and each segment has its own nerve ganglion situated on the ventrally positioned nerve cord. Embedded in pits actually within the ectoderm are chitinous bristles called chaetae. The earthworm has only | four chaetae per segment. It belongs to | the group of the annelid worms called the oligochaeta (Greek, olichos, few). Earthworms are the largest in size, and by far the most important of this group; the freshwater and_ shoredwelling oligochaetes, are all small and inconspicuous. Muchmore spectacular and diverse in appearance is the largest group of annelids — the polychaeta. Their name refers to the many chaetae they possess (Greek, poly, many). An alternative name for them is the bristleworms. Polychaete annelids are nearly all marine; in fact they are one of the most successful groups of marine invertebrates — possessing a divergence of structure and habit equalled only by molluscs and crustaceans. Unlike oligochaetes, polychaete worms are either male or female — they are not hermaphrodites. The third and last group of the annelids is the leeches or hirudinea. Relatively few in number, the leeches are highly specialized creatures, adapted to a parasitic mode of life. Polychaetes — Bristleworms The chaetae or bristles of the polychaetes are not only more numerous than in the oligochaetes. Instead of ~~ = ld al adhoetihoudy adh 4d AAO gn Ma scl Ok ES 2090 The Other Annelids being embedded in the skin, they arise in bunches from special leaflike prominencies of the body wall called parapodia (side-feet). Each segment possesses a pair, one on each side of the body. Parapodia are well-displayed in one of the commonest shore-dwelling polychaetes, the ragworm (Nereis). Each individual parapodium consists of an upper half, the notopodium and a lower half, the newropodium ; both support their own bunch of chaetae. Above and below each parapodium are flaps of skin, the dorsal and the ventral cirrus. Another feature in which polyCIRCULAR CROSS-SECTION THROUGH EARTHWORM (OLIGOCHAETE) CHAETAE Cross-sections of Earthworm, an oligochaete and Nereis

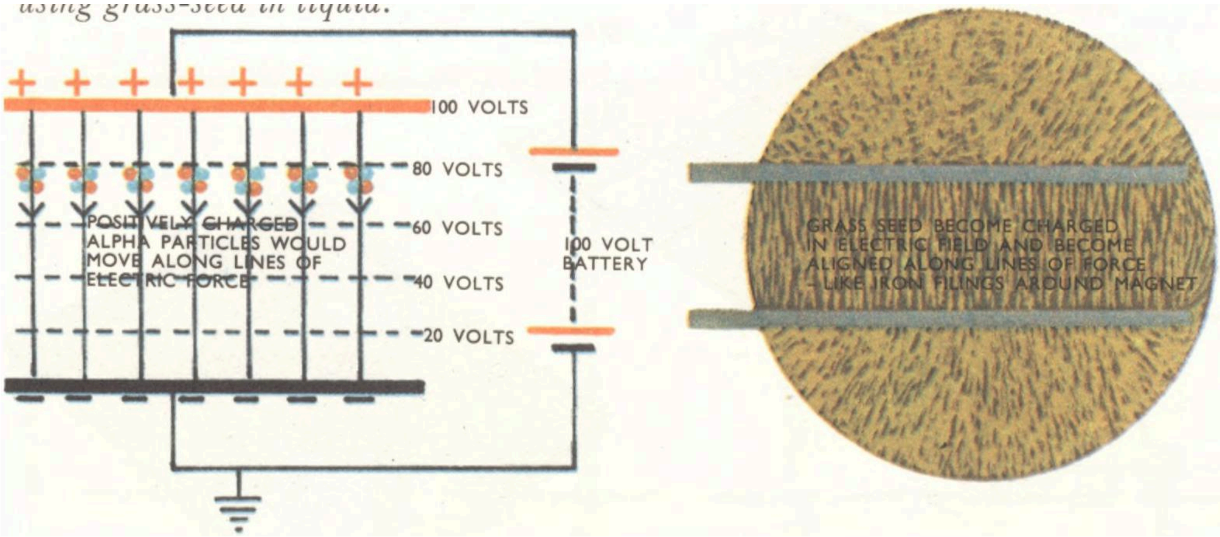
movements of the body cause water to flow in and out, carrying oxygen and possible indications of food in the vicinity. When small animals do pass the burrow, the first part of Nereis's gut (or pharynx) is turned inside out so that it extends from the burrow. On the end are two horny jaws for gripping the prey. Another group of wandering polychaetes are the scale worms. The upper surface of these creatures appear to be covered with armour plates, causing them to resemble the molluscan chitons. The plates are called elytra and in fact are adaptations of the CROSS-SECTION THROUGH NEREIS (POLYCHAETE) PARAPODIUM DORSAL CIRRUS VENTRAL CIRRUS a polychaete. Polychaetes are distinguished by their numerous chaetae and their side-feet or parapodia. Muscles activate the side-feet which can be used for locomotion. All sorts of variations occur in the shape and forms of chaete parapodia and cirri. chaetes differ from oligochaetes is the formation of a distinct head. Earthworms have a segment or prostomium in front of the mouth but there are no features suggesting a concentration of sensory organs. In Nereis, however, the prostomium is equipped with a pair of sensitive tentacles above and a pair of palps below. The next two segments have become fused together. Of all the segments of Nereis, they alone lack parapodia; but the cirri are still there — modified as sensitive tentacles. Nereis is an example of an errant polychaete, a worm which may wander using its rows of parapodia as paddles for swimming or just undulating its body. Actually most of its time is spent in shallow burrows. The sharp bristles which at other times probably protect the creature now enable the animal to grip the smooth walls of its burrow, while gentle undulatory The most modified of all polychaetes is Chaetopteris which lives within the sand in a U-shaped parchment-like tube. Some of the parapodia are large and fan-shaped ; they circulate water through the tube. Food particles are caught in a specially secreted mucus bag which is periodically swallowed. SY Sy , "e> 2: a yf: MUCUS > DR BAG SECRETED ,. yh BY EPIDERMIS = UA S4I08 uff, JAETOPTERIS A HIGHLY MODIFIED BURROWING WORM dorsal cirri; these structures have curved upwards and backwards on either side of the body and overlap on the top surface. Scale worms are usually small—rarely more than an inch or so in length. A notable exception however is the ‘seamouse’ (Aphrodite). Not at all like a worm in appearance this creature, which burrows in muds in off-shore waters, is 6 inches or so in length and two inches across. Its scales are not visible, for covering the upper surface is a dense coat of matted hairs. The hairs are really modified chaetae formed from the notopodium —the upper part of the parapodium. Short irridescent chaetae, projecting laterally, are used for locomotion. Safe in its burrow, water is continually pumped through the space between the back and the covering of hairs. The scales probably absorb oxygen. Nereis, the ragworm, has been desfor reproductive purposes, a remarkable change takes place both in its habits and its structures. As eggs and sperm develop in the bodies of female and male worms, the parapodia become larger and the spiny chaetae become flattened — almost oar-shaped. The colours of the worm (and they vary enormously with species) become richer and in the males eyes become enlarged. With the change complete the worms take to the open waters where eggs and sperm are liberated into the sea. This sexual stage of the Nereis is known as the heteronereis; formerly heteronereids were thought to be different animals altogether — and not surprisingly so. In the paddleworms, parapodia and dorsal cirri are paddle-shaped all the time and the animals swim as well as crawl. A remarkable pelagic worm is Tomopteris. Its parapodia have developed into enormous lobes. On the head are two very large ‘antennae’. In fact these are really parapodia much modified and supported by very long chaetae. Distinct from the wanderers and shallow burrowers are the tubemaking polychaetes — they actually construct their own homes. Appendages on the head are usually far more numerous than, for instance in Nereis, and greatly modified and increased in NEREIS - THE RAGWORM Nereis — the ragworm. The ‘rags’ are in fact the projecting parapodia. Inset,the head of cribed as primarily a burrower. But,) | ee eyes Ween: Saneory organs and an eversible, jawed pharynx. TENTACLE A terebellid worm. The colourful tentacles on the head are numerous and modified for capturing small food particles. The tube is normally buried in the sand. ELYTRUM (MODIFIED DORSAL CIRRUS) APHRODITE - THE SEA MOUSE Aphrodite, the Sea-Mouse is a scale worm. The dorsal cirri are turned backwards forming a protective coat. But the scales (elytra) tangl hairs — modified 'ANTENNA' are covered b chaetae. TOMOPTERIS —— Tomopteris — a pelagic polychaete. The parapodia used for swimming are enormous. The antennae are highly modified parapodia. size. Tube-living worms are filter feeders. The tentacles projecting from the head have ciliated grooves down which food particles are washed into the mouth. They do not have the eversible pharynx or jaw of Nereis and the parapodia are usually reduced in size. The serpulid tube-dwellers make their shells of calcium carbonate —a substance which they secrete themselves. Some tubes are relatively straight but others so tightly coiled that they resemble the spiral shells of snails. Another feature of sabellids is that the peristomium — the fused segments forming the head — is produced into a collar which folds back over the outside of the tube and secretes reinforcing, hoop-shaped rings. One of the tentacles from the head has a swollen branch; when the animal withdraws into its tube the structure fits exactly over the mouth of the tube. The sabellids make tubes by cementing sand and mud particles together The tubes are either buried in the sand or attached to rocks. Sabella pavonina — the peacock-worm — is a magnificent sight. Its tube is a foot long and from the opening, brightly coloured ciliated tentacles project. Terebellids make mud or sand tubes. In addition to food-collecting tentacles, they have 3 pairs of gills behind the head, for respiration. The serpulids and sabellids have no special structure but respire over the whole body surface. Of all polychaetes, the worm which most closely resembles the earthworm in appearance is Arenicola, the lugworm. The lugworm is a burrower and like the earthworm consumes vast quantities of mud from which organic food is obtained. Evidence of its activities are the conspicuous coiled worm casts on any sandy beach. Chaetae are small in the front and middle of the body, but absent altogether at the rear; the middle region, however, carries pairs of feathery gills on the dorsal surface. Sensory tentacles are absent; they would only be a hindrance to burrowing. But the pharynx can be extended. Instead of jaws the surface is coated with minute papillae which by adhesion enable the worm to pull itself forward through the sand. 2091

TRANSFORMER — SOURCE OF ALTERNATING POTENTIAL VARIABLE RESISTOR RI RANSFORMER Electric Fields HERE are three different sorts of forces that act on matter — gravitational, magnetic, and electric. We know from every-day experience that these invisible but all-important forces exist. When a magnet is separated from a piece of iron, when a heavy load is lifted, or when arubbed hair-comb lifts a piece of paper, magnetic, gravitational or electric forces are at work. Mystery still surrounds their true nature, but the scientist still needs to be able to detect and measure them. One thing that all forces have in common is that they must be described not only in magnitude but in direction. In other words, they are vector quantities. For example, it is not sufficient to say that a force of 1 pound weight, or 32:2 poundals, acts on a mass of 1 pound at the surface of the Earth. In ‘a. practical problem the structural engineer, for example, must put in a direction arrow to show that this force acts in a direction pointing towards the centre of the Earth. He knows that it acts in this direction because, if it were free to move, that is the way a beam or a girder would fall. 2092 Similarly, ifit were possible to separate off a free south magnetic pole of a magnet it would move towards the north pole of a nearby magnet. A magnetic force acts on the free magnetic pole in a definite direction. A free negatively-charged electron will move towards a_positively-charged plate because of the electric force acting upon it, also in a definite direction. Around the Earth, a magnet, or an accumulation of electric charge there is a region of influence in which a mass, a magnetic pole, or an ELECTRICITY The pattern of lines of potential inside and around complicated electrodes are plotted using an electrolytic trough. The potential at a point is found by inserting the probe and adjusting the resistance boxes (R, and Rz) until there is no sound in the head-set. From values of R, and Rg the potential is calculated. electric charge will ‘feel’ a force and in which it will move if it is allowed to do so. These regions of influence are called fields, and the paths that are traced out in them are called fteld lines or lines of force. The magnetic field lines around a magnet are easily demonstrated in the well-known method, using iron filings. They can also be plotted using a small magnetic compass. With a magnet, it is not possible to demonstrate field lines by the movement of a free single magnetic pole because a north pole can never be separated from the south pole in a magnet, but the result does show how a single magnetic pole would move if it were free to do so, ina single two-dimensional plane. Gravitational field lines may be demonstrated just as easily in a twodimensional plane. Ifa plane-board is tilted, a ball-bearing rolling down the plane from the top will trace out a path following a gravitational field line. This will show how the gravitational force is acting on the ball in the plane of the board, and repeated runs of the ball will trace out a field line 'map' on the board. Another 'map', of equal importance, may also be drawn on the board. The ball runs down the plane because at the top it has high potential energy and at the bottom it has low potential energy. It runs from a position of high potential energy to a Gravitational field lines. A ball bearing rolls down the inclined plane, having high potential energy at the top and low potential energy at the bottom. Parallel horizontal lines of equal potential may be drawn across the board. The field lines are traced out by the different paths of the ball along the board, at right angles to the potential lines.

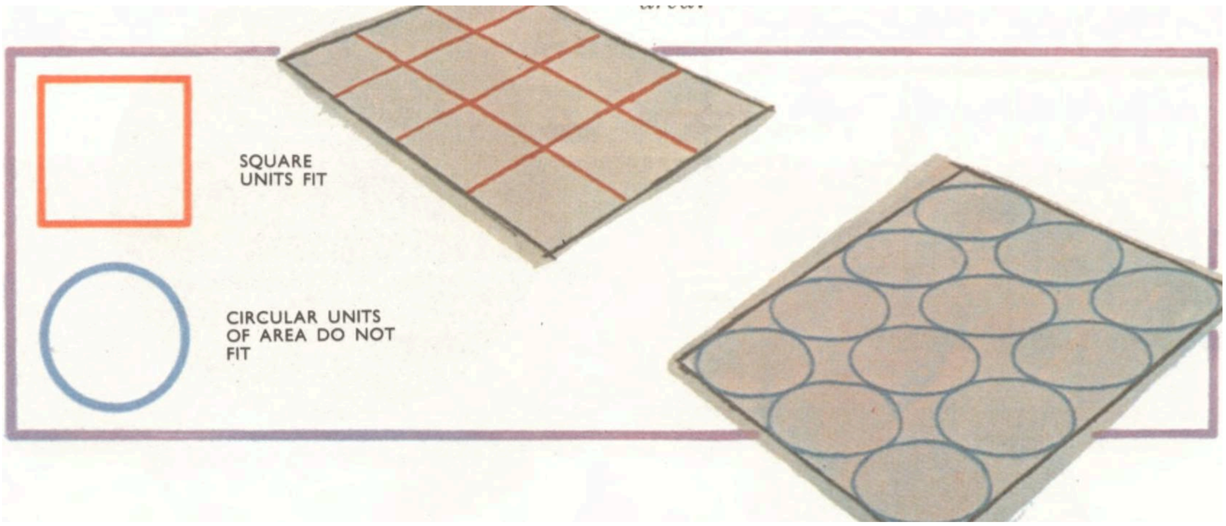

TRANSFORMER – SOURCE OF ALTERNATING POTENTIAL
VARIABLE RESISTOR R1
VARIABLE RESISTOR R2
AT PROBE, PROPORTION OF TOTAL TRANSFORMER POTENTIAL IS MATCHED WITH POTENTIAL IN TANK
NO SOUND IN EAR PHONES INDICATES 'BALANCE'
TRANSFORMER
PROBE
ELECTRODE SURFACES FLUSH WITH LIQUID SURFACE
RESISTANCE BOX R1
RESISTANCE BOX R2


EQUIPOTENTIAL LINES-BALL HAS HIGH POTENTIAL ENERGY AT TOP, LOW POTENTIAL ENERGY AT BOTTOM
FIELD LINES IN PLANE OF BOARD ARE 'DRAWN' BY PATH OF BALL
CIRCULAR POTENTIAL LINES ON CONE – FIELD LINES CUT THEM AT RIGHT ANGLES

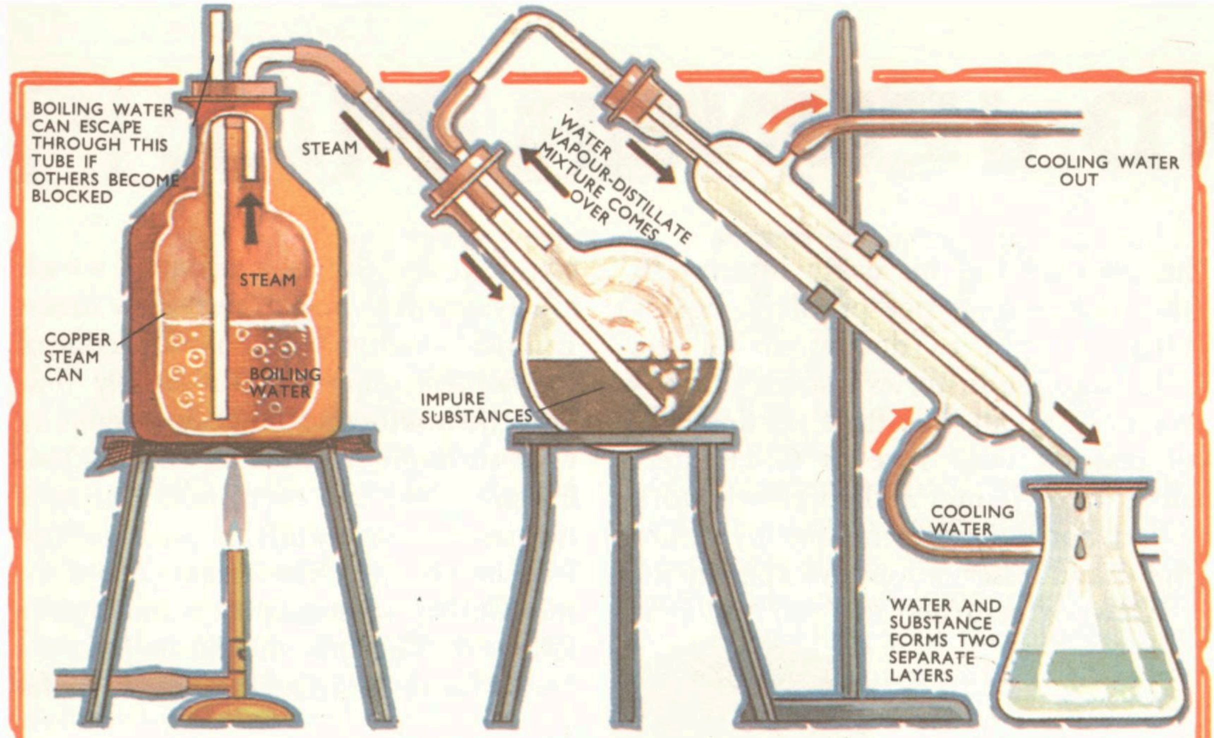
position of low potential energy. It is possible to draw lines across the board that connect points of equal potential energy. Similar lines of equal gravitational potential are useful in seeing what will happen in a three-dimensional example. Then the lines become contours, as on a relief map that shows mountains and valleys. If, for example, they were drawn on a perfect cone-shaped 'hillock' they would be a succession of circles of increasing radius, each one representing a contour of gravitational potential. On an irregular hillock the contours would be irregular, but the path of a ball-bearing could easily be predicted — it would run from a high potential contour to a low potential contour, and then to a contour of still lower potential. Further, it would run along the gravitational line of force that passes through each contour, crossing the contours at right angles. In an electric field a very similar thing happens. A pair of parallel metal plates is sealed into an evacuated chamber, one plate connected to the negative terminal of a battery, and the other to the positive terminal. The first plate becomes positively charged and the second plate negatively charged. The negative plate is at a lower electric potential than the positive plate so the positive plate is at the top of an electrical 'tilted plane'. If a positively charged particle (say an alpha-particle emitted by a radioactive element) is passed into the chamber it will move from the region of high electric potential to the region of lower electric potential — from the positive to the negative plate. In doing so, it will follow an electric field line and it will cross the contours of electric potential. Supposing the positive plate is at 100 volts and the negative plate is 'earthed'. Then, a 100 volts potential contour will pass through the positive plate and a succession of contours, running parallel with the plate, will run through the gap at perhaps go volts, 80 volts and so on until the o volts contour will be reached at the negative plate. The path of the alpha particle will cross each of the potential contours at right paths the electrons in, for example, an electron microscope are going to follow because the accurate focusing of the final image depends upon it. In general, the patterns of potential and field lines are very complex and difficult to find. Single electrons or alpha particles cannot be used to trace out a path because they are invisible. What is done is to construct models of the system and predict the behaviour of the electrons in the real equipment. In one such method, a rubber sheet is stretched over a number of props. Non uniform fields have curved lines which do not run parallel to each other. All these lines of force can be demonstrated using the ‘grass seed’ method (see illustration below) or may be found by plotting the potential lines in an electrolytic trough. ELECTRONS WOULD TRAVE ALONG LINES FROM NEGATI RADIAL LINES OF FORCE FROM CHARGED INSERTION OF METAL PLATE angles. At the central region between the plates the contours will be parallel, undistorted and evenly spaced. This is a uniform field and it is very difficult to achieve in practice because the edges of the plates have disturbing effects on the field. The plotting of electric fields in practice is a difficult process but it has to be carried out in the design of many pieces of electronic equipment. It is obviously important to know what The top plate is at 100 volts potential, the horizontal plate at zero potential. A positively charged particle would move along the field lines (lines of force) from positive to negative at right angles to the potential lines. (right) Field lines demonstrated between parallel plates using grass-seed in liquid. RADIAL LINES ETO AROUND CHARGED SPHERE D ‘AIR. OF CHARGED SPHERES The heights of the props are made proportional to say, the voltages on the different electrodes in the system so the sheet is stretched into a form that represents the variation of potential across the whole field of the equipment. Ball-bearings are then allowed to roll down the different ‘hills’ and their paths are similar to the paths that electrons would take up in the real equipment. Varying the heights of the props is equivalent to varying electrode potentials, and the effects of them may be studied,. as well as the effects of varying sizes and shapes of the electrodes. The potential distribution inside a \real piece of equipment may be found by inserting potential probes — metal rods of which the potentials may be measured. These, however, affect the distribution and shapes of field lines so a true answer cannot be found. This problem is overcome in another field model — the electrolytic trough. It is found that if model electrodes are placed in a conducting liquid then the readings taken using potential probes do not affect the field lines. 2093 *using grass-seed in liquid.*



MATHEMATICS A fraction of a person 1s only a statistician' s figmentation. People occur in whole numbers. COUNTING the number of people in a room is easy. One person = one unit. When dealing with whole objects it is easy to add and subtract because one object = one unit. However, very few convenient units of length occur in nature. Units like the foot and the yard do have some form of natural basis. A foot is roughly the length of a man's foot and a yard the length of a man's stride. The foot and the yard are not very well defined by this, because different people have different sized feet and varying strides. When making any measurement of The scientific length unit, the metre, ws defined as 1,552,734: 5 wavelengths of the red light emitted by a special kind of cadmium lamp. This arbitrary choice 1s nevertheless accurately reproducable in scientific laboratories throughout the world. 2094 There are no obvious natural units for length. Man has had to invent purely arbitrary units. T length, area or volume, the measurement must always be made in standard, well-defined units. Since there are no obvious natural units for any of these quantities, Man has devised convenient artificial standards — such as the foot and the metre. Units of this size are useful for measuring the height of a chair, or the length of a room. But no-one would quote the diameter of a microscopic bacterium in metres— they use far smaller units, the micron, or the Angstrom unit. These smaller units are in fact based on the metre. A micron is a millionth of a metre, and an Angstrom unit is a ten-thousandmillionth of a metre. Small units based on the foot are never used in scientific work. Larger distances are also measured in larger units—the mile, and the kilometre. Distances to stars are so SQUARE UNITS FIT CIRCULAR UNITS OF AREA DO NOT FIT Length Units Factors of 12, 3, and 1,760 are involved in the British System, of inches, feet and yards. 10 is the only factor needed to convert from smaller to bigger units in the metric system. (m = metre) | | Angstrom unit (A) = 10,000,000,000 m | | Micron (1) T;000.000 m | millimetre (km) 80 m Icentimetre(cm) = | 100 m | TOm 10m 100m 1,000 m | decimetre (dm) | decametre | hectometre (hm) | kilometre (km vast that they are quoted in light years. This unit is based neither on the foot nor on the metre. It is the distance light can travel (at 186,000 miles' per second) in a year. One light year is 5,880,000,000,000 miles. Units of Area An area of carpet can be measured in square feet, square yards, or, in the metric system, in square metres. The unit of area is the square, the side of which is unit length. The area of a circle is also quoted in ‘square units’. It would, of course, be possible to have as a unit the area of a circle of unit radius. But it would be very difficult to measure an area with a measuring circle. The only shape the circle would fit on to would be a circle of equal radius. Square units are easily fitted round each other. In the same way, the unit of volume is the cube of unit radius — the cubic inch, the cubic yard (for measuring cement) the cubic centimetre, the cubic metre. Squares are better than circles as units of area.

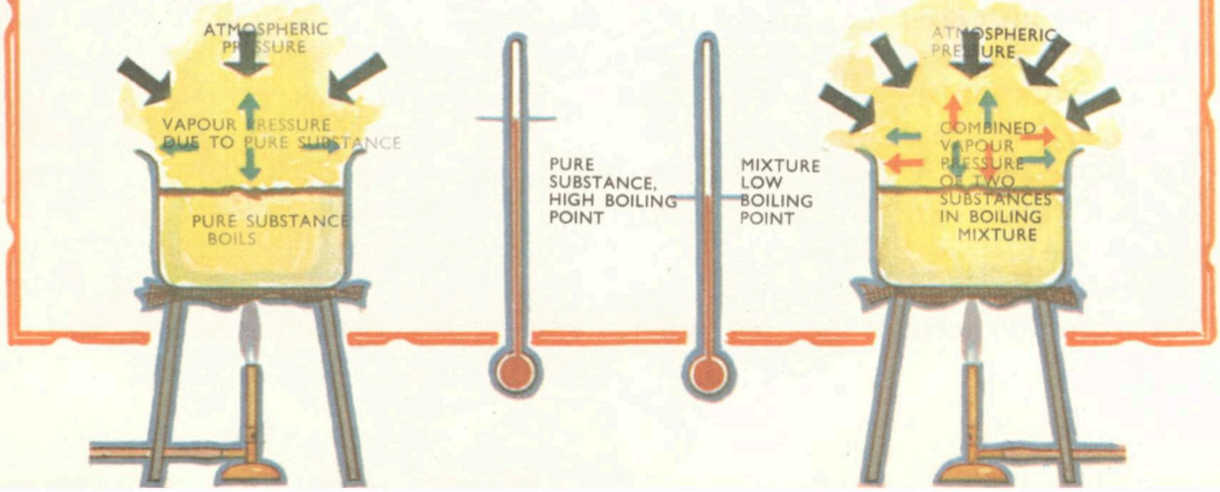


COPPER STEAM IMPURE SUBSTANCES COOLING WATER OUT ‘WATER AND SUBSTANCE FORMS TWO SEPARATE LAYERS If a substance has a high boiling point, but does not mix with water, it can be distilled over at'a temperature below | °C by bubbling steam through it. Steam Distillation ORGANIC chemicals with very high boiling points can often be purified by steam distillation. Organic substances are rarely made pure in one single process. Usually the first preparation results in a mixture containing a host of unwanted substances which have to be removed at a later stage. Distillation often provides the answer. But ordinary distillation has certain drawbacks. Crude chlorobenzene, (CgH;Cl), with its many tarry impurities, presents this problem. Its boiling point is high (132°C), and therefore it must be heated strongly for distillation to begin. Some of the impurities start to char when it is strongly heated. Instead of boiling and bubbling gently, the contents start erupting and bumping and the far-too-hot flask could be cracked by a particularly violent bump. Instead of heating the distillation flask with a Bunsen burner, the flask can be successfully heated by bubbling steam through its contents. The steam keeps the mixture well stirred and at a temperature of around go°C, a watery mixture of chlorobenzene distills off, leaving the impurities behind. The chlorobenzene is insoluble in water and the two quickly separate out into two easily separated layers The crux of the matter is the insolubility of the chlorobenzene in Pure water boils when its vapour pressure, increased by raising the temperature, equals atmospheric pressure. When chlorobenzene is added, the vapour pressure of the chlorobenzene is added to the vapour pressure of the water. The total vapour pressure is then increased with a rise in temperature and the mixture boils at a lower temperature than the pure water. PURE SUBSTANCE, HIGH BOILIN POINT MIXTURE LOW OILING POINT PHYSICAL CHEMISTRY water. Steam distillation will only work for substances that are virtually insoluble in water. For when two substances are immiscible in one another they are capable of quite independent behaviour. Water and chlorobenzene are immiscible. Water on its own, when it is heated, evaporates more and more until at 100°C the pressure of its vapour has risen to that of its surroundings. and the water boils. Chlorobenzene on its own behaves similarly, but boils at the higher temperature of 132°C. At a particular temperature, both pure liquids exert their own particular vapour pressures. Mixing chlorobenzene with water will not alter these boiling points, for the two substances behave independently. At a particular temperature, the vapour pressure of the mixture can be calculated by adding the vapour How much chlorobenzene comes over with the water? This depends on the partial pressure (p.p.) at the boiling point and the molecular weights (m.w.) mass of chlorobenzene mass of water MW chlorobenzene m.w. P-P- chlorobenzene x P-P- water water When the calculation is carried out (for an external pressure of 740 mm. of mercury) it is found that 71% of the distillate, by weight, should be chlorobenzene. This figure is verified in practice, so the steam distillation process is remarkably effective in bringing over the chlorobenzene at a temperature much lower than _ its boiling point. pressure of the water to that of the chlorobenzene. So the mixture being heated more quickly arrives at the pressure of its surroundings than would either pure chlorobenzene or pure water. And the boiling point of the mixture is therefore lower than that of pure chlorobenzene or pure water. Steamy chlorobenzene is driven over into the condenser. \s the total vapour pressure in the listillation flask does not depend on the amount of steam or chlorobenzene present, the chlorobenzene will continue to distill off at the same temperature as long as there is any of it present. 2095



If a substance has a high boiling point, but does not mix with water, it can be distilled over at a temperature below 100°C by bubbling steam through it.

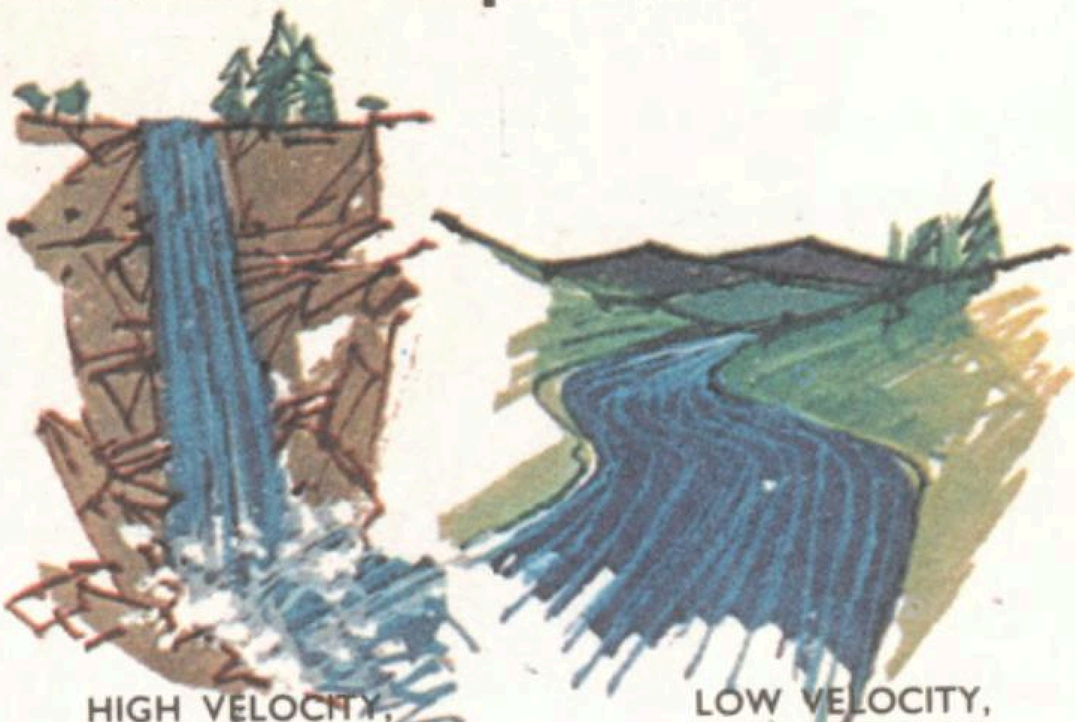
Pure water boils when its vapour pressure, increased by raising the temperature, equals atmospheric pressure. When chlorobenzene is added, the vapour pressure of the chlorobenzene is added to the vapour pressure of the water. The total vapour pressure is then increased with a rise in temperature and the mixture boils at a lower temperature than the pure water.



FAMOUS SCIENTISTS EDWARD JENNER, M.D. WHEN he died in 1823, Edward Jenner was a celebrated international figure. Grants in recognition of his work had been awarded to him by the British and Indian Governments, an honorary degree was bestowed upon him by the University of Oxford, and his name was familiar throughout Europe, America, and Asia. Yet for much of his life, Jenner had lived and worked in obscurity. Gloucestershire was Jenner's native country. He was born in the village of Berkeley on May 17th 1749, the.son of a clergyman of the district. During his early education at nearby Woottonunder-Edge and Cirencester, he developed a keen interest in natural history, particularly geology and ornithology. Already he showed the enquiring spirit and powers of observation which were to serve him so well. Jenner decided upon a medical career and accordingly he became an apprentice to a surgeon in Sodbury, near Bristol. Three years later he journeyed to London, where his new teacher was the famous 18th century surgeon John Hunter. Hunter was an inspiring figure and encouraged Jenner in both his medical and natural history studies. The habits of the cuckoo, hibernation in animals, and the life history of eels were just a few of the subjects the two men investigated. One of Jenner's early interests — one in fact which occupied him whilst still an apprentice at Sodbury — was the means of preventing outbreaks of smallpox. Smallpox in the 18th century was still one of Man's worst scourges. Periodically epidemics would break out killing thousands of people. Local tradition in Gloucestershire had it that those people who suffered from cowpox— a mild and harmless disease — would remain unaffected by the fatal smallpox. There were many stories to support the tradition and Jenner pondered deeply. Though an able musician and poet, and a great favourite of society, Jenner cared little for city life. In 1773, he returned to Berkeley to practise medicine. Encouraged by Hunter, 2096 he also revived his early interest in the smallpox-cowpox problem, beginning in 1775 to document all the relevant case histories known. Unfortunately about this time occurrences of cowpox were rare in Gloucestershire and Jenner's theoretical work could not be supplemented by either first-hand observations or experiment. Cowpox returned in 1796 and Jenner acted promptly. Material from a cowpox vesicle on the hand of a local dairy maid was taken and injected into the blood stream of an 8-year-old boy — one James Phipps. Two months later, Jenner performed the crucial experiment. He injected virulent smallpox material into the boy. And the boy remained well. The experiment was repeated — again with success —and in 1798 Jenner wrote a memorable paper on the subject. Chance injections had occasionally been tried in the past but Jenner's careful work put the subject of inoculation against smallpox onto a firm scientific basis. A stage had been reached when an all-out crusade against the disease could begin. Jenner's immediate confidence had not taken into account the unfounded criticism and strong resistance which always seem to accompany new ideas. Eminent doctors with no knowledge of the subject objected. But finally Jenner's own self-confidence and tenacity won through. Returning to London, he prepared his own material and inoculated thousands of poor people free of charge. Significant drops in the fatality of smallpox immediately followed. The tide turned in Jenner's favour and in 1802 he received the One of Man's greatest benefactors, Edward Jenner. By his work, smallpox was conquered and the way was cleared for the prevention of many other diseases. first of his grants from the British Government. In the following year the Jennerian Society was founded to promote widescale inoculation against smallpox. The Society flourished until 1808 when the National Inoculation Establishment was founded. Rightly, Jenner became a celebrated man throughout the world. Honours showered upon him. But with characteristic disdain for fame and fortune Jenner returned to his post of country doctor at Berkeley. He died in 1823. Jenner had the satisfaction of seeing smallpox vanquished. But the value of his work did not end there. The principle of vaccination was eventually to lead to the immunization of Man against other dangerous diseases.

APPLIED SCIENCE | PROPELLERS and CAVITATION At the tip of a large high-speed ship's propeller blade, water is forced to move backwards at a high velocity. What effect has this on the water ? One important and inevitable effect is that the pressure in the water drops. The higher the water's velocity, the lower the pressure becomes. This is as predicted by Bernouilli's Equation, the basic equation of fluid dynamics. When the pressure drops, the boiling point of the water drops. With a reduced pressure pushing them back into the body of the liquid, more and more water molecules are able to escape into the vapour state. A vapour bubble starts to form in the liquid, and quickly grows as more vapour molecules escape into it. This effect is known as cavitation. So, at high propeller speeds, the water starts to boil around the leading edge of the propeller. Bubbles of vapour form in the water. Some of these implode near the surface of the propeller blade with sufficient force to blast off small pieces of metal, and eventually eroded. This large transatlantic liner that the propellers had to be replaced after her maiden voyage. With better-designed propellers, the problem is not so acute, and even lowspeed vessels can be operated with a small amount of cavitation biting away at the edges of the blades. But cavitation, a form of turbulence, leads to an extra drag on the propeller blades, and causes a drop in the overall efficiency of the propulsion system. There are no half-measures with cavitation. Propellers can be operated with none at all (or a very small amount), or a very large amount — full cavitation or super-cavitation. No propeller would last long if it were operated with a middling amount of cavitation and it would be very inefficient, so this region of operation is avoided. The jump to full cavitation is made possible by specially-designed propellers, with the flat aerofoil crosssections reminiscent of supersonic aircraft and very sharp leading edges to bite into the water. The sharp edges are essential to reduce the crosssectional area affected by cavitation, and hence reduce the extra drag caused by this effect. The pressures are so low under full cavitation that the bubbles implode at some distance from the surface of the blades, where they BUBBLES IMPLODE Y M BUBBLES IMPLODE AWAY FROM PROPELLER Bernouilli's Equation The main conclusion drawn from Bernouilli's equation is that the pressure ina fluid drops when its velocity increases. This turns out to be another statement of the Law of Conservation of Energy. The pressure in the fluid is unlikely to remain the same — it can either increase or decrease when the fluid velocity increases. At first sight, an increase looks more likely, but it would violate the Law of Conservation of Energy. An increase in pressure would lead to an increase in velocity, a further increase in pressure and the process would rapidly get out of hand. can do no damage. These high-speed propellers are used for hydrofoil craft, which cruise at around 60 knots. Submarines are also capable of travelling at high, undisclosed speeds, but they normally travel underwater. Here there is little danger of cavitation. The pressure in the water increases with depth. Underwater, the pressure includes the weight of the water above, as well as atmospheric pressure. Although the pressure is lowered in the vicinity of the submarine propellers, it can never drop low enough for the water to boil and cavitation to set in. LOW PRESSURE, CAVITATION REGION 2. BUBBLES FORM Above: a section across an ordinary ship's propeller, with its almost-aerofoil cross-section. propeller designed for full cavitation. Below: the sharply-pointed CAVITATION ACROSS THIS ENTIRE SURFACE < ; FLUID FLOW 2097

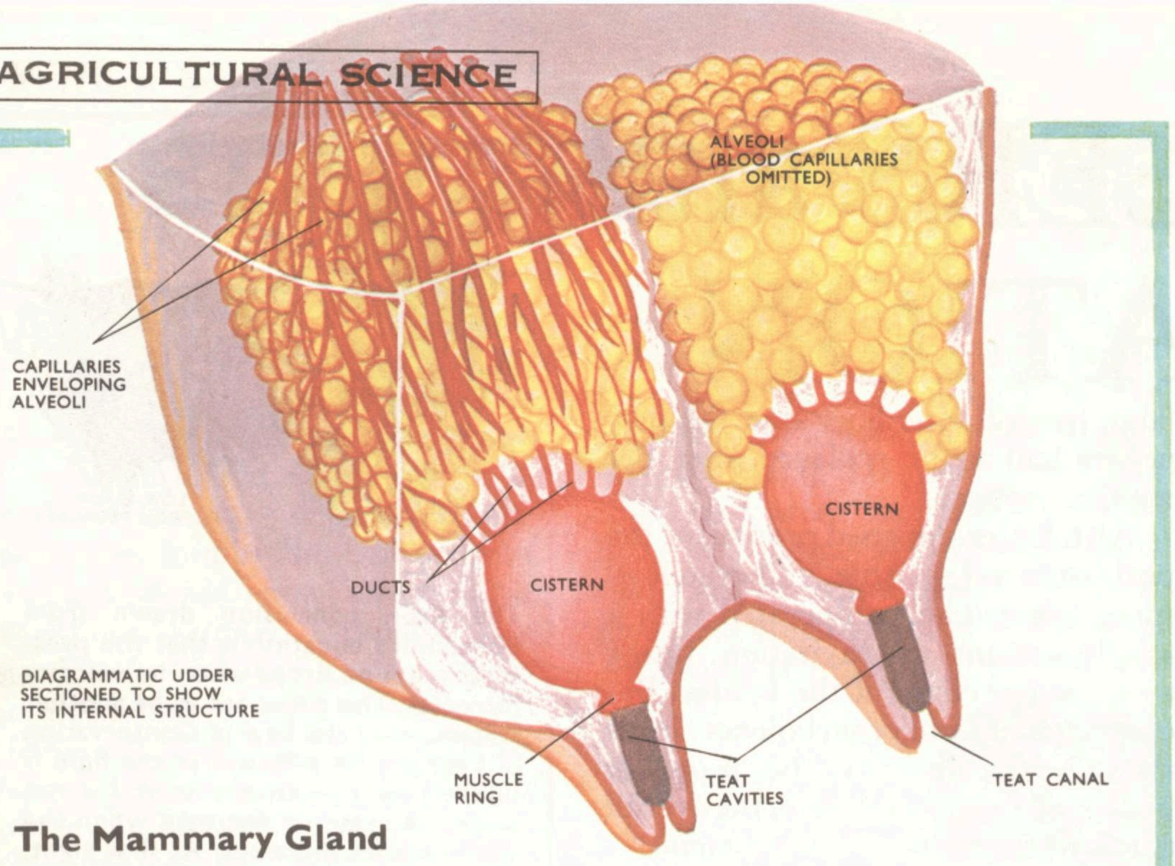
Bernouilli's Equation



The main conclusion drawn from Bernouilli's equation is that the pressure in a fluid drops when its velocity increases. This turns out to be another statement of the Law of Conservation of Energy. The pressure in the fluid is unlikely to remain the same – it can either increase or decrease when the fluid velocity increases. At first sight, an increase looks more likely, but it would violate the Law of Conservation of Energy. An increase in pressure of Energy. An increase in pressure would lead to an increase in velocity, a further increase in pressure and the process would rapidly get out of hand.

CAPILLARIES ENVELOPING ALVEOLI DIAGRAMMATIC UDDER SECTIONED TO SHOW ITS INTERNAL STRUCTURE MUSCLE RING The Mammary Gland TEAT CANAL A cow's udder consists mainly of clusters of gland cells grouped to form numerous tiny, hollow sacs or alveoli. Each alveolus is surrounded by blood capillaries from which the gland cells obtain their nourishment and the substances that they convert into milk. In the lactating animal milk is poured continuously into the hollow centre of each alveolus and leaves by a duct. The ducts of the many alveoli join like the tributaries of ariver and empty their contents into the storage chamber or cistern above each teat. It is thought that the mammary glands may be modified sweat glands because of certain similarities in structure. The BIOLOGY of MILK MILK is one of the richest foods available to man. Apart from eggs, it is the best source of animal protein and would appear to provide a ready answer to the world's food problem of protein deficiency. A pint of cow's milk contains almost nineteen grams of protein and supplies nearly four hundred Calories. In Israel the average milk yield per cow has reached the phenomenal figure of 4,000 kilograms (around 650 gallons) per year — a protein yield of around 95 kilograms. It is probable that domestic cattle are descended from the wild aurochs — Bos primigenius, the bones of which are common in Pleistocene deposits — which were living until the 17th century. The probable area of domestication was the Middle East, some six thousand years ago. Although dairy farming was practised in Mesopotamia as long ago as 3000 B.c., no record of dairying in Europe has been 2098 found before 1000 B.c. Since that time various kinds of cattle have been introduced into Britain by the Romans, Saxons and Norsemen and also from the Channel Islands. From this hotchpotch of material the breeds that we are familiar with today have been developed. Most of the basic work in establishing the breed characteristics of today's cattle was carried out in the eighteenth and nineteenth centuries. What is Milk? Milk is the fluid produced by the mammary glands, after the birth of the young. It is a fluid characteristic of mammals; these are the only animals to suckle their young. From a commercial point of view cattle are by far the most important producers of milk and references to milk apply to cattle, unless otherwise qualified. The metabolism of a dairy cow is 'geared' to producing milk which is why her build is much slighter than that of a cow reared for beef.

Milk is composed principally of water — around 87 per cent — and also contains proteins, milk sugar or lactose, most vitamins — even though the quantities of some are negligible traces — fat and various salts. It is particularly rich in calcium, which is of obvious importance to the growing young animal whose sole source of food initially is milk. Calcium is required in large quantities for bone. formation. The amount of calcium in human milk is, not surprisingly, much lower than that in cow's milk. The demands of a growing calf are way beyond those of a much slower growing and smaller human youngster. The percentage of protein in cow's milk is correspondingly higher. Milk is a type of colloid known as an emulsion (a 3-4°% emulsion of fat stabilized by milk protein), that is, physically it consists of a liquid dispersed within a liquid. Its ivory coloration is a result of the almost equal scattering of all wavelengths of light. The slightly yellow tint of the cream is due to the pigment carotene (this is the precursor of vitamin A — important in vision — which is formed in the liver). A layer of cream forms on top of the more watery remainder because the larger fat droplets float to the surface. Milk is usually slightly acid, having a pH of between 6-6 and 6:8. The main protein of milk is casein (or caseinogen). About three and a half per cent of milk is protein: two thirds of this is casein. The enzyme, rennin, acts on Casein Causing it to clot. It is especially active in young mammals whose sole diet consists of milk. Diagrams showing the major constituents of cow's milk (below) and human milk (below right). The principal constituent 1s water. Note that the percentage of protein in cow's milk is higher than inhuman milk asisthe salt



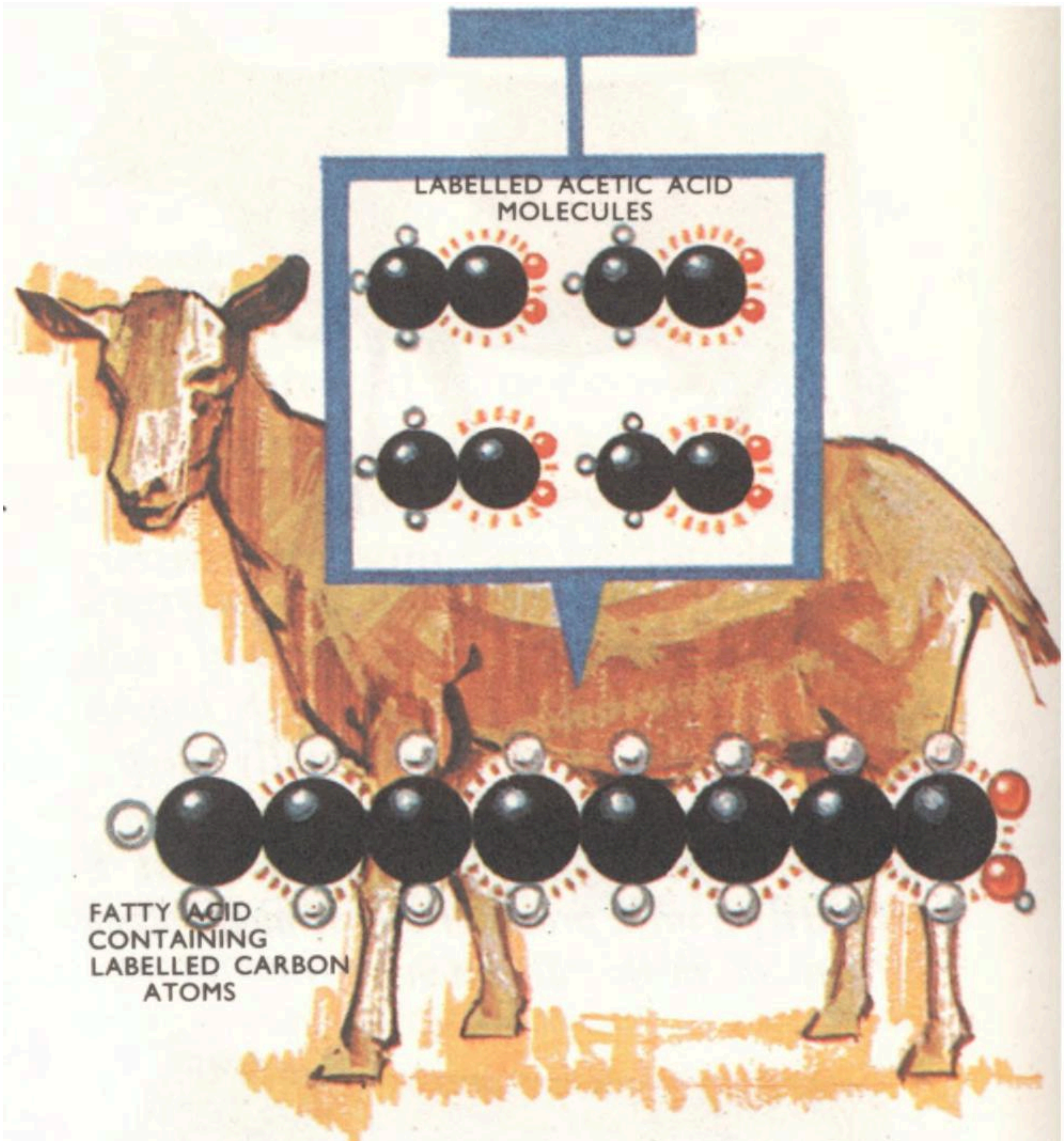
The Mammary Gland

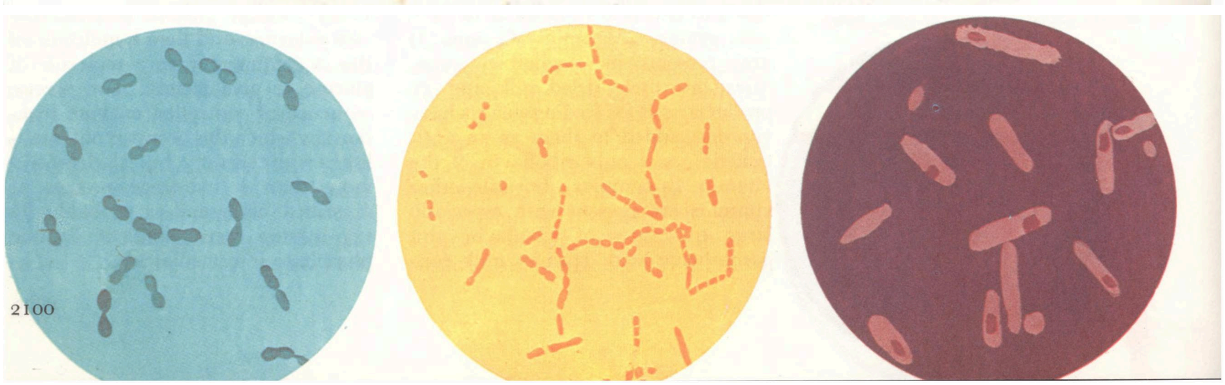
A cow's udder consists mainly of clusters of gland cells grouped to form numerous tiny, hollow sacs or *alveoli*. Each alveolus is surrounded by blood capillaries from which the gland cells obtain their nourishment and the substances that they convert into milk. In the lactating animal milk is poured continuously into the hollow centre of each alveolus and leaves by a duct. The ducts of the many alveoli join like the tributaries of a river and empty their contents into the storage chamber or *cistern* above each teat. It is thought that the mammary glands may be modified sweat glands because of certain similarities in structure.

Illustrations showing the characteristw builds of dairy cattle (above) and beef cattle (below). The latter have a much heavier build and the chest is especially deep— metabolism is diverted towards flesh production. The lighter built dairy beast ws primarily a milk producer whose metabolism is 'geared' in this direction. ABERDEEN ANGUS-BEEF BREED Analysis of the casein molecules shows it to contain all the essential amino acids that the body is unable to make itself. The only sugar present in milk is lactose. It is found in the milk of all mammals and is formed in the mammary gland and no other part of the body. Lactose is a disaccharide, its molecule being composed of one molecule of glucose and one molecule of galactose. Just under five per cent of lactose is usually present in cow's milk, but human milk contains up to seven per cent and a cow elephant as much as seven and a half per cent. In a doe rabbit only two per cent lactose is found. Thus the composition and quality of the milk of different species proportion, the latter being largely due to the large quantities of calcium present. The growth rate of a calf is much greater than that of a child hence the greater requirement Sor protein (muscles) and calcium (bones). of mammals varies considerably. It is interesting that in whales there is 12 times more milk-fat and four times more protein than there is in cow's milk. Fat is of great importance as a heat insulatory material to aquatic mammals and also as an energy store for their long migrations from polar seas to more temperate waters. Souring of milk is due to the fermenting activities of Streptococcus lactis and many other organisms infecting it. Enzymes that these microorganisms contain convert lactose into lactic acid. The increase in acidity precipitates the protein in milk producing the characteristic curdling. The principal site of fat synthesis in the vertebrate body is the liver. But, in mammals, the mammary glands are important sites of fat production outside the liver. Much of the body fat is composed of long chain fatty acids but milk fats contain considerably more short chain fatty acids such as butyric acid and caproic acid. This is probably related to the great production of organic acids in the rumen (a chamber of the stomach) from the cellulose of plant material by the actions of microorganisms. There is much less butyric acid and caproic acid in human milk where bacterial action in the gut takes place on a much smaller scale. The principal fats are triglycerides such as tripalmitin, tristearin, and triolein. In addition, milk contains lipids such as lecithin and cholesterol. The latter is related chemically to the bile salts and the sex hormones. The main salts present in milk are the chlorides and phosphates of calcium, sodium and potassium. Calcium and phosphorus are particularly necessary for teeth and bone formation. Each litre of milk contains 1,200 milligrams of calcium (cheese, incidentally, contains nearly that amount in each 100 grams). Vitamins C and D are present in negligible quantities (this is why dried milk, and the products used for feeding infants have vitamins added to them as an antiricketic), and only riboflavin of the vitamin B complex, but all other vitamins are present with especially large quantities of riboflavin and pantothenic acid. Human milk contains more ascorbic acid. The Formation of Milk The udder of a cow has a rich blood supply and prominent milk veins can be seen on the abdomen ofa dairy cow. It might be expected that the milkproducing cells of the udder merely filter substances from the blood, thus making milk. But this is only partly the case. Certainly blood contains sugars, fats, proteins, vitamins, salts and all the other ingredients of milk. However, the sugar in the blood is glucose: that of milk is lactose. Obviously, therefore, chemical changes occur in the udder during the formation of milk, and although, the complete chemistry of its formation is far from being known, the little that we do know confirms the existence of a delicate, yet elaborate, metabolic mechanism. The larger fat droplets present in milk separate out on top of the watery remainder forming the cream. The yellowish colouration is due to the pigment carotene. Most of our knowledge has been derived from the use of isotopic tracers. For example, labelling of glucose with carbon-14 — a radioactive isotope of carbon—has shown conclusively that milk sugar (lactose) is formed from glucose derived from the blood. Firstly, glucose is converted into galactose and then a molecule of this is combined with a molecule of glucose to give lactose. The precise sequence of controlled enzyme reactions by which this occurs is not established with certainty, but it is definitely more than a two-stage process as described here, and is probably by step-by-step conversions of organic phosphate intermediates.

Similar experiments 'labelling' fat in the blood have shown that some of the milk fat is derived from blood fat without chemical change. The most important source of fatty acids is probably acetic acid formed in the rumen by the breakdown of cellulose by micro-organisms. Each acetic acid molecule contains two carbon atoms. By means of enzymes these two carbon units are linked together forming chains containing as many as sixteen atoms of carbon. Palmitic acid has sixteen carbon atoms in its molecule. But fatty acid molecules containing eighteen or more carbon atoms, (e.g. stearic acid), are almost certainly obtained 'ready-made' from the blood. The paths of protein synthesis are not known. It is fairly certain that proteins from the blood are not used to make caesinogen. Its molecule is built up from the 'free' amino acids in the blood flowing through the mammary glands. The Control of Milk Secretion There appears to be no actual control of the mammary gland by another distant part of the body as far as milk quality is concerned. The actual quality of milk produced depends to a certain extent on the cow's diet, though the latter can vary enormously without effect. A general deficiency of food lowers the volume of milk produced but not the quality since the cow's reserves are drawn upon to make up for dietary deficiencies. An example of this is the drain on calcium reserves. A diet continuously low in calcium results in a great loss of calcium from the bones, in acute cases, even producing skeletal abnormalities which may interfere with the calf-bearing abilities of the cow. The butter fat content of milk may vary, however, with the diet. A cow fed on young grass or on a diet low in roughage but high in concentrates may show a butter fat depreciation of more than fifty per cent. This is largely due to the fact that microbial activity is much reduced in the rumen since the bacteria have less material to ferment. When a cow is milked there is a delay of about half a minute before 'let down' — the time when the milk begins to flow freely. Prior to 'let down' there is a build up in the pressure in the ducts of the mammary gland. This is due to the action of the hormone oxytocin on the muscle cells surrounding the ducts and alveoli of the gland. Oxytocin is produced by the hind part of the pituitary gland, in response to massage because of the resulting nervous reflex. It is carried in the blood to the mammary gland where it causes the muscle cells to contract — ultimately with the ejection of milk through the teats. (The same hormone causes the musculature of the uterus to contract during the birth of a calf—

and a human child). The actual formation of milk is probably initiated by another hormone, prolactin, produced by the anterior lobe of the pituitary gland. The quantity of milk produced depends to a large extent on the suckling action of a calf or the equivalent action of milking. Frequent and complete milking is thus an important part of dairy cattle management. Suckling and mechanical milking, through nerve Se Vo Ve Ve Ve Ve Injection of labelled acetic acid into goats resulted in short chain fatty acid being built up from the 2-c acetic acid units. Labelled fatty acids were later isolated from the milk. reflexes, seem to cause increased production of prolactin which therefore stimulates the milk producing cells to greater activity. The development of the mammary gland in the young cow and the proliferation of its tissues during pregnancy are controlled by a host of hormones — principally sex hormones such as oestrogen and progesterone. Their action in fact prepares the mammary gland for the action of prolactin. A cow does not produce milk until it has had a calf, and in order to obtain as much milk as possible from a cow during its lifetime it is desirable for it to have several calves and thus several periods of lactation. Some of the bacteria that occur in milk or milk products. From left to night Streptococcus lactis an organism that causes milk to sour by producing lactic acid, Streptococcus thermophilus and Lactobacil tridium tyrobutyricum an anaerobic bacterium causing blowing of hard cheeses. us bulgaricus used in the making of yoghourt and Clos


LABELLED ACETIC ACID MOLECULES
FATTY ACID CONTAINING LABELLED CARBON ATOMS



ELECTRONICS Aerials—Resistance to Radiation SIGNAL RADIATED AWAY FROM THE DIPOLE IN ALL DIRECTIONS The radiation resistance 1s a measure of the power which can be transmitted away from a half-wave dipole, working at tts resonant frequency. An alternating signal of the current frequency does not notice the gap at the centre of the dipole, or the termination of the circuit. It travels along the wire just as if a 73 ohm resistance were there in place of the gap. "TRANSMITTING aerials are similar in principle to receiving aerials. They are based on the halfwave dipole, a piece of conducting metal of length half the wave-length of the transmitted signal. The dipole is divided at its mid-point into two parts. One electrical lead goes to each part. The signal is in the form of to-andfro surges of current (alternating current). The surges can be transferred around the circuit, from one half of the dipole, through its connecting lead, and along the other lead to the other half of the dipole. The gap in the middle of the dipole, which would stop any flow of direct, one-way current, does not prevent the alternating surges from flowing, to-and-fro, along the halves of the aerial. But, as in all alternating current circuits, the taming of the current surges must be right for the maximum amount of current to flow to-and-fro. Part of a surge flowing back down an aerial can easily cancel out the following part of the surge flowing up the aerial. However, when the dipole halves are quarter of a wavelength long, the timing is right. The circuit is said to be tuned. Surges up and down the aerial reinforce each other at all points. The current flowing in the aerial is a maximum. The signal can then travel to an aerial just as easily as if a 73 ohms resistance were at the end instead of the dipole aerial. 73 ohms is the radiation resistance, and it is virtually the same for all half-wave dipoles. - If a 73 ohms resistance were connected in place of the dipole aerial, it would dissipate the same amount of power as the aerial radiates into space. The radiation resistance is an equivalent resistance — it is not a real, physical resistance — and it can be calculated from the power radiated away from the aerial, and the current at the centre of the aerial. Whenever currents change in circuits, part of the energy of the electric current is radiated away as electromagnetic radiation. An unchanging direct current does not radiate. Some Steady currents radiate none of their energy, but whenever currents change, some energy is radiated. Parallel wires leading to the aerial tend to radiate, but the radiations cancel out. acceleration in the current is necessary for this to happen. Currents are accelerating in the two leads to the transmitting aerial, but these do not radiate either. The reason is that the two leads are parallel to each other. Both tend to radiate, but the radiations are equal and opposite, and cancel each other out. Two kinds of leads are commonly used for aerials. In one, the two leads are spaced apart from each other and parallel. The other common form is a coaxial cable. Here, one lead surrounds the other, preventing the lead picking up unwanted signals from outside. The coaxial cable joined to a dipole aerial is equivalent to a resistance of around 73 ohms. This is called its characteristic impedance. Coaxial cables could be made with a variety of characteristic impedances, but 73 ohms is used with a dipole aerial because it 'matches' the radiation resistance of the aerial. This is the condition for the maximum transfer of power along the circuit. FAD CURRENT second mx ot fee RADIATION CHANGING CURRENT 2101

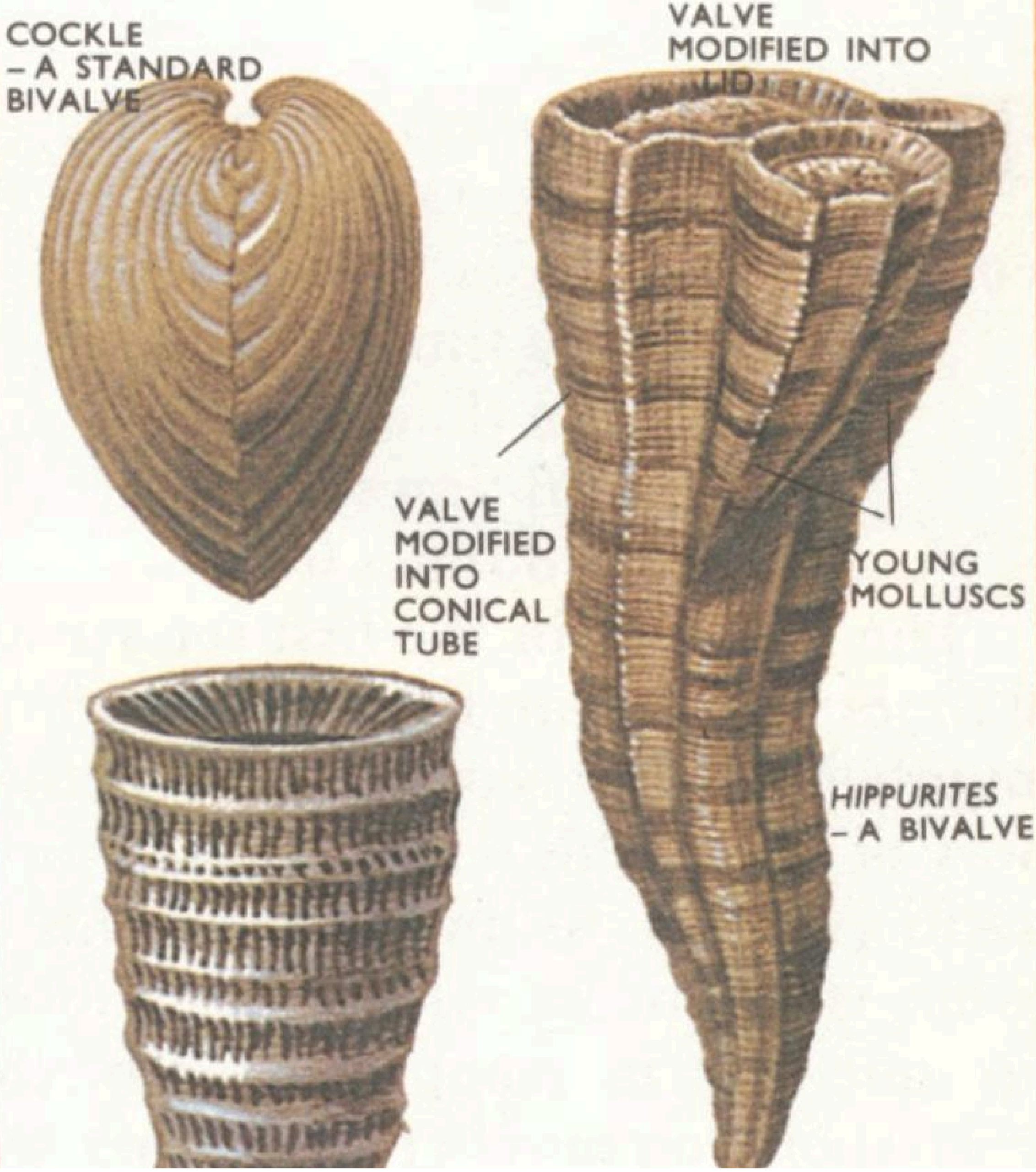
BIOLOGY CONVERGENT AND PARALLEL EVOLUTION F whales, Hermann Melville, in his great sea story 'Moby Dick', wrote, 'To be short, then, a whale is a Spouting fish with a horizontal tail'. Melville, a seaman and the acutest of observers, was not a trained biologist. Whales do of course resemble fish. They are streamlined in shape. They have a fluked tail, a pair of flippers as fore limbs, and they may even have a dorsal fin along their backs. But in fact whales are mammals; they have warm blood and the females suckle their young with milk. Thus, despite all appearances, they are far more closely related to dogs, rabbits and Man himself than any fish. One hundred and fifty million years ago another fish-like creature swam the seas. Though it became extinct 70 million years ago, fossils of its bones enable its appearance to be reconstructed. But, far from being a fish, the preserved skeletons show beyond all doubt that the creature was a reptile. Because of its remarkable resemblance to a fish it was called an ichthyosaur or fish-lizard (Greek, ichthys, a fish; sauros, a lizard). Mammals, reptiles, fishes — three distinct vertebrate groups —but all with members resembling one another. The connecting link between whale, fish and ichthyosaur is that all have lived in the sea. All have become adapted to an A placental and marsupial mole. Long snouts and powerful claws give these two burrowing animals a remarkably similar appearance. PLACENTAL MOLE MARSUPIAL MOLE aqueous mode of life. Evolving in the seas, similar types of structure have developed. This evolutionary development, which often leads to apparent likenesses between stocks that have diverged a long way from an original ancestor, is called convergent evolution or simply convergence. Convergence is a common phenomenon found in organisms living on land, in water and in the air. It is a powerful witness to the theory of evolution by natural selection. Under the same selective pressures different groups will respond with apparently similar adaptations. Care, then, has to be often taken in classifying animals and plants. Superficial resemblances between two organisms do not necessarily mean that they are closely related. More Examples of Convergence Cut offin the continent of Australia, the marsupials or pouched mammals have been able to flourish for 150 million years. In most other parts of the world they have disappeared, for competition from the placental (nonpouched) mammals proved too much. Australia offers a wealth of environments, and many different marsupials evolved, filling the niches. The similarity in appearance between the various marsupials and_ placental mammals provide striking examples of convergence. Thus Australia has its wolf-— the Tasmanian ‘devil’; it has its native ‘cat’ (Dasyums) and its native ‘mouse’ (Dasycercus). There are ‘ant-eaters’ and there are ‘sloths’. The flying phalanger is comparable with the flying squirrels and the wombat with ground hogs. Australia even has its own marsupial mole. Marsupials and placentals are all mammals and do possess a common ancestry in the not-too-remote past. Convergence can, however, produce similarity between completely unrelated creatures. Thus insects and BEE HUMMING-BIRD birds are far removed from one another, yet at a glance it is very difficult to tell the difference between a humming bird and a humming moth. Both are the same in size, and both live off nectar in flowers and have converged in their hovering flight and _ their feeding procedures. Another remarkable example is the resemblance between the bill of the duck—a bird—and the bill of the platypus — a mammal. Both creatures obtain food by sifting mud; to perform this function they both have a broad flattened duckbill, giving their heads a similarity in appearance. Convergent animals may look alike but it is easy to show that they are entirely different creatures with very unlike ancestors — their resemblance in appearance is not due to close relationship. The structures which give the resemblance, often do not develop from a common feature in an ancestor. Thus, the wings of the humming moth are utterly unlike the wings of the humming bird in their origins. Such structures with unrelated origins but with the same functions are said to be analogous. In parallel evolution the very same features possessed by the common ancestor may evolve in very similar ways in the descendants. The two structures, because of this correspondence, are said to be homologous. Homologous organs always indicate some common ancestry.

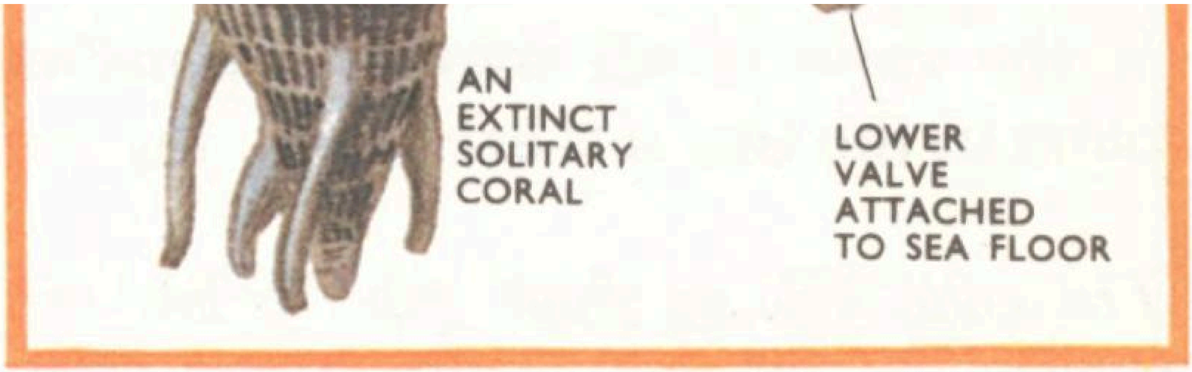
BEE HUMMING-BIRD

HUMMING MOTH Birds and insects — two groups very far removed from any possible common ancestor. But the similarity of their modes of life has given the humming birds and humming moths a close superficial resemblance. Of burrowing creatures, probably the best known is the common earthworm — a long slender creature with no notable structures projecting from its surface. Animals from other groups that have taken up a burrowing mode of life often possess an identical shape. For instance, there are the slow-worms and skinks (really lizards) and the caecilians (really amphibians). The shape and general appearance of the snakes also suggests that they evolved from a burrowing group. Convergence is not confined to the animal kingdom. Plants exposed to similar outside influences may also resemble one another in the structures they evolve. Hundreds of species of tree make up the three great rain forests of the world. Yet even to a specialist it is often difficult to tell one form from a completely unrelated form. All grow to great heights, lack branches on their lower trunks, develop similarly shaped foliage and similar bark. In American deserts cacti are characteristic plants. Their stems are swollen with water-storing tissues; they are covered in protective spines and do not carry leaves. Very similar in all these respects are the euphorbias, native to Africa. But despite the cactuslike appearance, euphorbias are not closely related to cacti. Parallel Evolution The term parallel evolution or parallelism also describes a process of similar Ichthyosaur (reptile) and whale (mammal) and fish — three types of swimming vertebrate. The fore-limbs are all really homologous, showing that the three groups have an ancestry, however remote, in common. The groups have since become widely separated and similarity in appearances, 1s only a consequence of convergent evolution. WHALE ICHTHYOSAUR ee FISH adaptations under the control of natural selection. In convergence, however, ancestries are very different, as close inspection soon reveals. Marsupials are easily distinguished from placentals not only by their pouches but numerous differences in their skeletons. In parallelism ancestries are not so different. Species evolving from the same ancestor may, instead of steadily increasing their differences, evolve independently along the same lines. The jerboas of African and Asian deserts are rodents especially adapted to their environment. Particularly conspicuous are their long hind legs Hippurites —the fossil remains of a coral-like animal. In fact it is a bivalve mollusc.

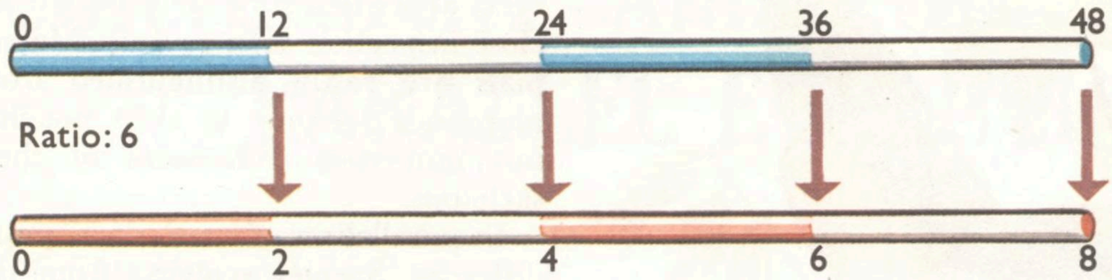
The two valves have become fantastically altered suiting a coral-like mode of life. VALVE MODIFIED INTO ATTACHED TO SEA FLOOR with which they leap rapidly over the sand. In America another rodent — the kangaroo rat — has almost exactly the same appearance, though it must have evolved entirely separately from the jerboa. Another example of parallel evolution may be seen in the porcupines of South America and those of Africa. In all probability, porcupines evolved from a spineless common ancestor and have developed their pines independently, in the two continents. 2103

Hippurites – the fossil remains of a coral-like animal. In fact it is a bivalve mollusc. The two valves have become fantastically altered suiting a coral-like mode of life.



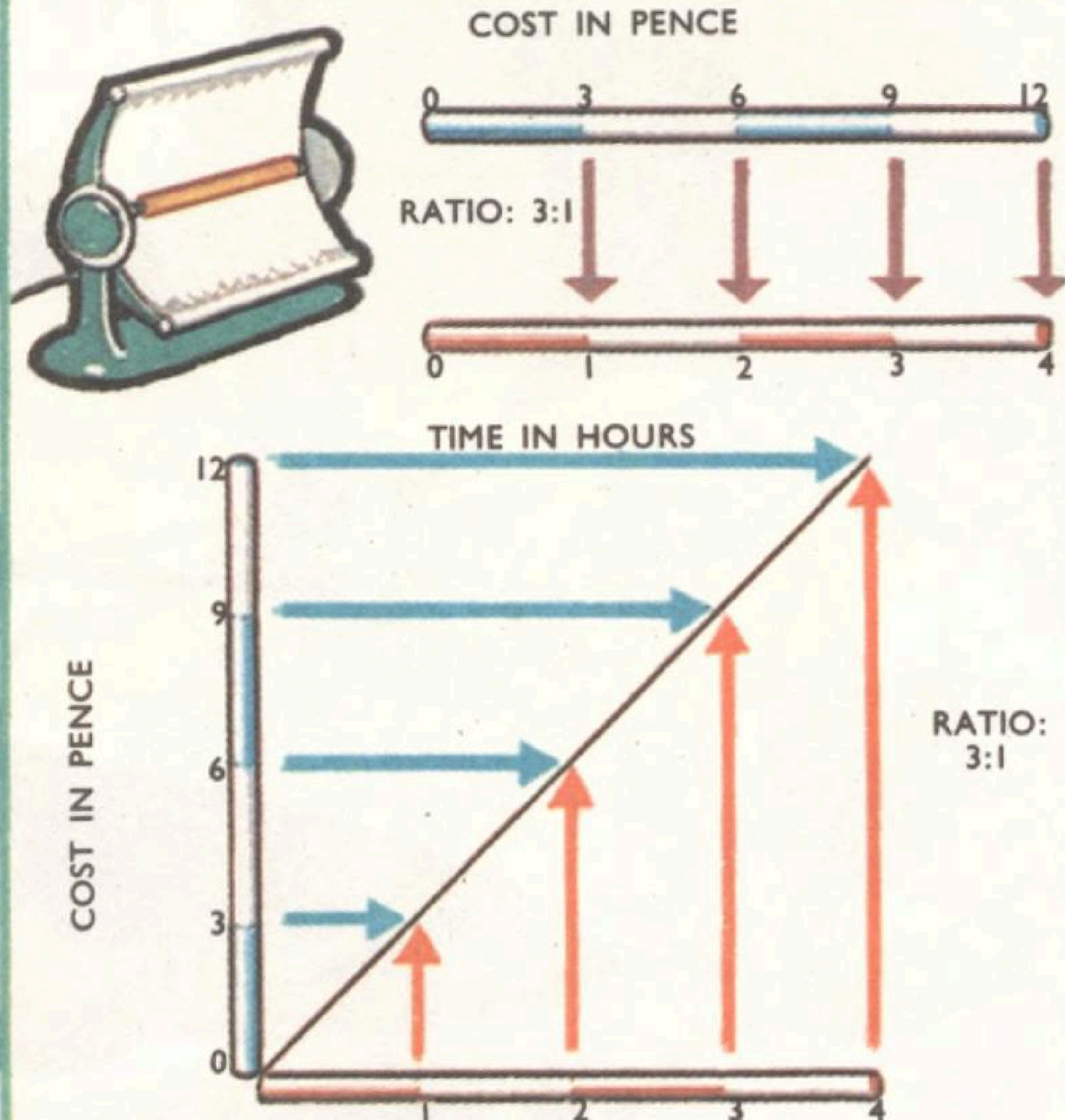

AN EXTINCT SOLITARY CORAL
LOWER VALVE ATTACHED TO SEA FLOOR

MATHEMATICS Ratio: 6 0 ¥ | Proportional Sets and Constant Ratio If two sets of numbers are proportional, they can be arranged on two different number lines so that all the members of one set correspond to members of the other set. The ratio (number in first set) : (number in second set) is constant. Ratio and Proportion "THE set of numbers 12, 24, 36, 48 has an important property in common with the set 2, 4, 6, 8. 12 is 2 times 6 24 is 4 times 6 36 is 6 times 6 48 is 8 times 6 The numbers in the first set are all 6 times bigger than the corresponding numbers in the second set. Another way of saying this is: ratio (12,2) = 6 ratio (24,4) = 6, and so on. This is what is meant by ratio — one number in one set is compared with the corresponding number in another set. If the ratio of all the corresponding pairs in the two sets are equal, the sets The usual kind of graph paper ts like two number lines at right angles to each other. Proportional sets give a straight line. 2104 are said to be proportional. Many pairs of quantities in real life are proportional to each other. If you leave an electric fire on for one hour, it will probably use up 3d. worth of electricity. After 2 hours, 6d. worth has been consumed. The cost of the electricity corresponds to one set of numbers 3 6 ae and the time in hours corresponds to another set I 2 3 4 Ratio (3,1) is 3. Ratio (6,2) is 3, and so on. These two sets of numbers are therefore proportional. It stands to reason that this must be right, for if two quantities are proportional, the bigger one is, the bigger the other is. The longer the fire is left on, the greater the total cost of the electricity. Cost is proportional to time. Another example is a man walking at a steady rate of 4 m.p.h. Provided he sticks to this steady rate, the distance he travels will be proportional to the time taken. At the end of the first hour, he has covered 4 miles, at the end of the second, 8 miles, and so on. The two quantities, distance and time, can correspond to two sets of numbers: distance in miles 4 8 12 16 total time in hours It 2 3 4 Again, it is obvious that distance is w Vv 4 rrr) a z a ° oO The cost of using an electric fire is proportional to the time for which it is used. One number line is used for cost in pence, the other for time in hours. In this example the ratio cost: time is 3:1 — the cost per hour of the electricity is 3d. per hour. The cost . at any intermediate time can be found from the graph. proportional to time. The ratio of distance in miles to total time in hours for corresponding terms in the set is always the same ~—in this example, 4. Once it is known that two sets are proportional, it is easy to find an unknown term in either of them. Suppose, for example, the problem is to find the distance travelled after 6 hours. To solve the problem, it must be mapped into mathematical symbols. The unknown distance corresponds to the unknown, x. x must belong to the set of distances.



Proportional Sets and Constant Ratio

If two sets of numbers are proportional, they can be arranged on two different number lines so that all the members of one set correspond to members of the other set. The ratio (number in first set) : (number in second set) is constant.

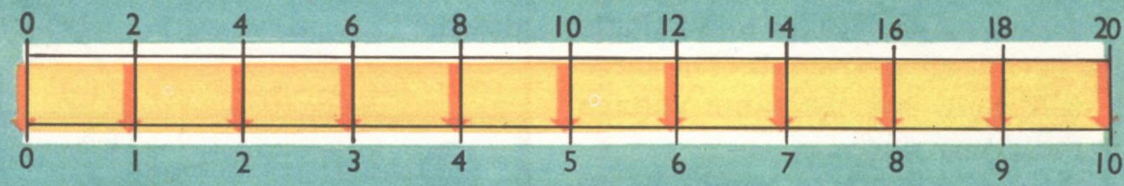


The cost of using an electric fire is proportional to the time for which it is used. One number line is used for cost in pence, the other for time in hours. In this example the ratio cost: time is 3:1 – the cost per hour of the electricity is 3d. per hour. The cost at any intermediate time can be found from the graph.

It must also be connected times bya ratio equation. This information is sufficient to pinpoint the value of x. x = 4 x 6 (multiply both sides by 6) X= 24 Solution: {24} —a singular set. Answer to problem: the distance travelled is 24 miles. The properties of proportional series can be put in another way. Equal steps in one set correspond to equal steps in the other set. In the example 1% (24 30048 I 2 3 equal gaps of twelve units in the upper set correspond to equal gaps of one unit in the lower set. This property is best seen when the two sets are drawn against each other in a graph. Problem | Proportional Sets and Straight Line Graphs . Equal parts of the set on the horizontal number line correspond to equal parts on the vertical line. The ratio (vertical step): (horizontal step) is constant. Another name for it is the gradient of the straight line. Right: The great advantage of graphs. It is possible to see whether two sets are proportional, whatever scale is chosen for the two number lines. In one diagram, two inches correspond to 6 units on the vertical number line. The scale in the other is one inch to 6 units. Both give the same ratio between the set of 6:1. x ‘ ‘ SSC See Se OTR ETE CS OE Ae eS eo, Problem 2 In an electrical circuit, the ratio (voltage:current) is known as the electrical resistance. If, in one particular circuit, it is known that the current is | amp when the voltage is 24 volts, and 2 amps when the voltage is 48 volts, what is the value of the current when the voltage is 30 volts? Ratio (voltage:current) is (24:1) and this ratio is the same for any particular circuit. If the unknown current is symbolized by x, then the ratio equation gives Ratio (30:x) = Ratio (24:1) The ratio is like an ordered pair of numbers, and the ratio equation can be solved by multiplying throughout by x and dividing by 24, to isolate x on one side of the equation. This gives: x = (30,24) — x = (5,4) — x= ltor 1-25 Answer: | -25. ; Solution to problem: the current is |-25 amps when the voltage is 30 volts. The volume of a gas is related to its temperature. If the volume of a certain amount of gas is 100 c.c. at 0°C, and 200 c.c. at 273°C, what is the volume of the same amount of gas at 100°C? The pressure is kept constant throughout. At constant pressure, the volume of a gas is proportional to its Absolute Temperature (°A), its temperature in °C plus 273°. 0°C is 273°A, 273°C is 546°A, 100°C is 373°A. Ratio (volume:Absolute Temperature) is (100:273) and, at constant pressure and with the same amount of gas, this ratioremains the same. If the unknown volume is symbolized by x, then the ratio equation is: Ratio (x:373) =

(100:273) 100 x 373 273

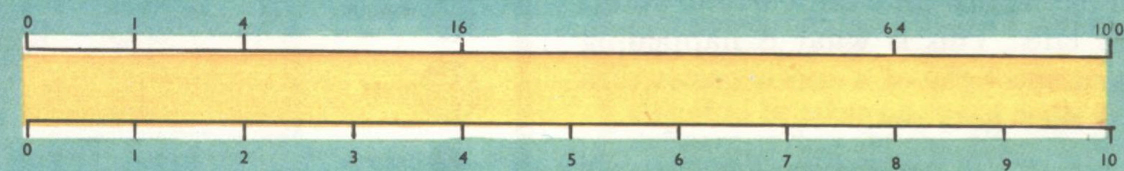
These sets are proportional



Proportional sets can be arranged along two number lines (with the scales suitably adjusted) so that there is proportionality all along the line.

Ratio {natural numbers which are multiples of two}: {natural numbers} is 2:1

These sets are not proportional



The upper set is {squares of natural numbers}. The lower set is again {natural numbers}. The ratio is not constant.

Problem 1

In an electrical circuit, the ratio (voltage:current) is known as the electrical *resistance*. If, in one particular circuit, it is known that the current is 1 amp when the voltage is 24 volts, and 2 amps when the voltage is 48 volts, what is the value of the current when the voltage is 30 volts?

Ratio (voltage:current) is (24:1)

and this ratio is the same for any particular circuit. If the unknown current is symbolized by x, then the ratio equation gives

Ratio (30:x) = Ratio (24:1)

The ratio is like an ordered pair of numbers, and the ratio equation can be solved by multiplying throughout by x and dividing by 24, to isolate x on one side of the equation. This gives:

$$x = (30,24)$$
$$\leftrightarrow \ x = (5,4)$$
$$\leftrightarrow \ x = 1\tfrac{1}{4} \text{ or } 1\cdot 25$$

Answer: 1·25.

Solution to problem: the current is 1·25 amps when the voltage is 30 volts.

Problem 2

The volume of a gas is related to its temperature. If the volume of a certain amount of gas is 100 c.c. at 0°C, and 200 c.c. at 273°C, what is the volume of the same amount of gas at 100°C? The pressure is kept constant throughout.

At constant pressure, the volume of a gas is proportional to its *Absolute Temperature* (°A), its temperature in °C plus 273°. 0°C is 273°A, 273°C is 546°A, 100°C is 373°A.

Ratio (volume:Absolute Temperature) is (100:273) and, at constant pressure and with the same amount of gas, this ratio remains the same. If the unknown volume is symbolized by x, then the ratio equation is:

Ratio (x:373) = (100:273)

$$\leftrightarrow \ x = \frac{100 \times 373}{273}$$
$$\leftrightarrow \ x = 136\cdot 6 \text{ approximately.}$$

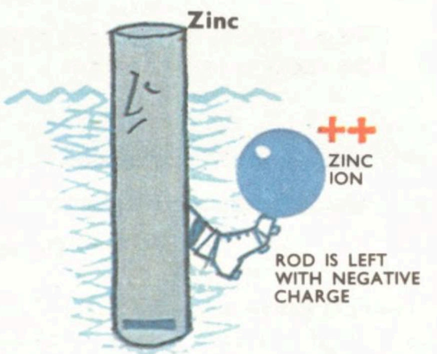
Answer: 136·6.

Solution to problem: the volume is 136·6 c.c. when the temperature is 100°C.

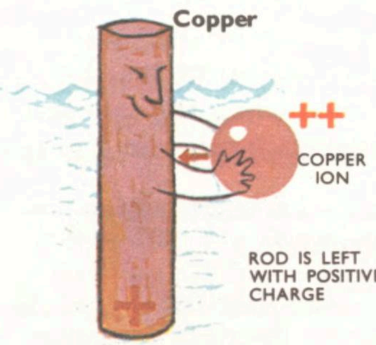
INORGANIC CHEMISTRY SUMMARY of CELLS JN an electric cell, chemical processes take place that create a surplus of electrons at one electrode and a shortage at the other. When the cell is connected to an external circuit, electrons flow from one electrode to the other, in attempting to ‘even up’ the electron inbalance. Each individual type of cell has its own starting voltage. A Daniell cell has a voltage of 1-1 volts and a dry cell (the sort used in torches) has a voltage of 1-5 volts. This applies to all single cells, both large and small. The large cell will have just the same voltage as a miniature one. The only advantage would be in useful life. The large cell with its large supply of chemicals would last a long while whereas a miniature one would be quickly used up. When any one of the necessary chemicals runs out, the voltage drops and the current falters and fails. This is what is happening when the light of a torch grows dim and then goes out. The Daniell cell and the dry cell are both primary cells. They use up chemical energy and give out electri¢al energy. When they run down there are two alternatives — fill up with more chemicals, or throw them away and buy another. There is no question of recharging by feeding in a direct supply of electricity to rejuvenate the cell. Recharging in this manner is for secondary cells. When they run down, more electricity can be passed in to charge them up again. This is the reason that the lead accumulator, a secondary cell, is used in motor vehicles in preference to the dry cell which would constantly need replacing. To provide the higher voltages used in practice, it is customary to place a number of cells in series so that their individual voltages add up. Two dry cells placed end to end in a torch will give an overall voltage of 3 volts — typical of bicycle lamps. 2106 When several cells are connected up in this way, the whole arrangement is known as a battery. A 224 volt battery for a transistor radio consists of 15 1°5 volt dry cells. Research is constantly being carried out to find new, more compact and efficient dry cells for use in transistor sets and guided missiles, for example. Also it is now possible to buy dry, portable secondary batteries that can be re-charged after they have run down. ZINC —0-51 VOLTS @ IRON @ HYDROGEN MORE ELECTRO=0:20 VOLTS POSITIVE +0-24 VOLTS COPPER Dprcrcow Gwe a CARBON +0-60 VOLTS +0-99 VOLTS LESS ELECTROPOSITIVE +1-06 VOLTS ABOUT | VOLT ROD IS LEFT WITH NEGATIVE CHARGE Zinc is strongly electropositive and tends to dissolve and form zinc ions. (Zn*). In solution, the positive zinc ions leave the zinc with a slight negative charge. After a time equilibrium is reached when the potential difference between the zinc and its surroundings is about —0-5| volts. ELECTRONS FLOW TO BALANCE UP CHARGES ON The cellacts"as a store of chemical energy which, when an external circuit is completed, will release energy in the form of an electrical current. Why a Cell Works The working of a chemical cell is based on the fact that some metals are more strongly electropositive than others. A metal with a strong tendency to go into solution as positively charged ions is a strongly electropositive metal. With other metals, the reverse is true and instead of dissolving a little, they tend to gather about them positively charged ions. One rod of each type dipping into an electrotype will make a cell of sorts. ROD IS LEFT WITH POSITIVE CHARGE Copper is not nearly so electropositive as zinc. In fact, it has the tendency to gather up positive charge from the surrounding electrolyte. At equilibrium the potential difference between the copper and its surroundings is +0-60 volts. When these two rods are dipped into the same dish of electrolyte and connected externally by a piece of wire, the electrons on the zinc rod can flow —— the wire to the copper rod to try to make up for the electron deficiency there. The voltage of the cell is |- 1 (0-5+0-6) volts.

Why a Cell Works

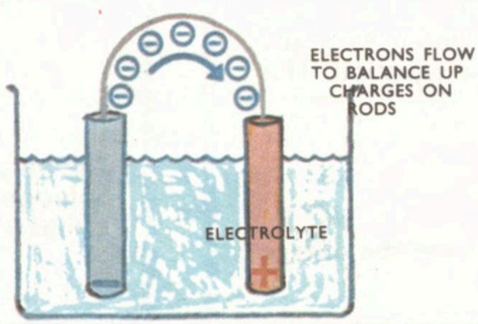
The working of a chemical cell is based on the fact that some metals are more strongly *electropositive* than others. A metal with a strong tendency to go into solution as positively charged ions is a strongly electropositive metal. With other metals, the reverse is true and instead of dissolving a little, they tend to gather about them positively charged ions. One rod of each type dipping into an electrotype will make a cell of sorts.



Zinc is strongly electropositive and tends to dissolve and form zinc ions. (Zn^{++}). In solution, the positive zinc ions leave the zinc with a slight negative charge. After a time equilibrium is reached when the potential difference between the zinc and its surroundings is about −0·51 volts.



Copper is not nearly so electropositive as zinc. In fact, it has the tendency to gather up positive charge from the surrounding electrolyte. At equilibrium the potential difference between the copper and its surroundings is +0·60 volts.

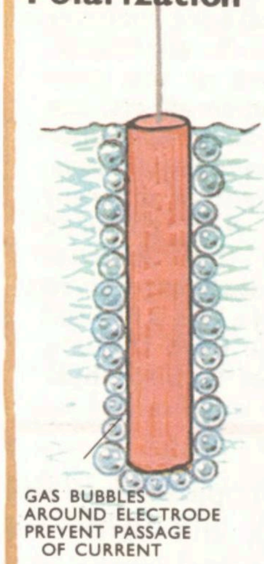


When these two rods are dipped into the same dish of electrolyte and connected externally by a piece of wire, the electrons on the zinc rod can flow through the wire to the copper rod to try to make up for the electron deficiency there. The voltage of the cell is 1·1 (0·5+0·6) volts.

Any electrode that attracts _ positive charge will attract hyd Sf ek ions which form bubbles of hydrogen gas, clouding the electrode and preventing the cell from working. To prevent this polarization, the cell must have some means of chemically removing the hydrogen ions before they have a AROUND ELECTRODE PREVENT PASSAGE arizin as es. OF CURRENT & & bubb Secondary Cells chance to form polDaniell Cell The Daniell Cell can maintain asmall current, delivered at I-1 volts, fora long period of time. DURING DISCHARGE LEAD (Pb) IONS COMBINE WITH SULPHATE IONS, DEPOSITING LEAD SULPHATE ON THE ELECTRODES This cell, the lead accumulator, used for car batteries, discharges in a similar fashion to the a cells. But the difference is that the chemical reactions taking place are reversible. nder discharge, the electrodes gradually turn to lead sulphate. When electrons are fed in the opposite way during charging the reverse reactions take place. The sulphate ions removed from the acid are put back and the electrodes become lead and lead dioxide once more. Alkaline Cells A cell or battery need not necessarily have an acid as electrolyte. Alkalis can also be used. A typical alkaline battery is composed of nickel-cadmium alkaline cells. This cell, like the lead accumulator, is a secondary cell, using up chemical energy while discharging and replenishing it under charge. A single cell has a voltage of about | -2 volts. New sorts of dry cells are being developed all the time. The mercury alkaline primary cell is capableaf doing the work of a dry Leclanché cell six times t Cadmium is more strongly electropositive than nickel and therefore ionizes, leaving an excess of electrons which flow through the external circuit to be taken up by nickel ions. Leclanché ZINC ROD MMONIUM HLORIDE SOLUTIO} MANGANESE BRASS CAP DIOXIDE AND POWDERED CARBON CARBON ROD (POSITIVE) AMMONIUM CHLORIDE PASTE MANGANESE DIOXIDE IS DEPOLARIZER The Leclanché cell, now rarely used, was the fore-runner to the dry cell used in torch batteries. The zinc is more electropositive than the carbon. The electrons left behind when the zinc ionizes flow through the external circuit to be taken up by the electrondeficient carbon. In the dry cell, the zinc is also the container. This cell is not used to supply electrical current-—it is a caomium Supue Standard against SULPHATE (77) which other volCRYSTALS VES] tages can be MERCUROU oe>ij accurately deaehare ney termined. The PASTE cell voltage of a Weston cell is ZINC

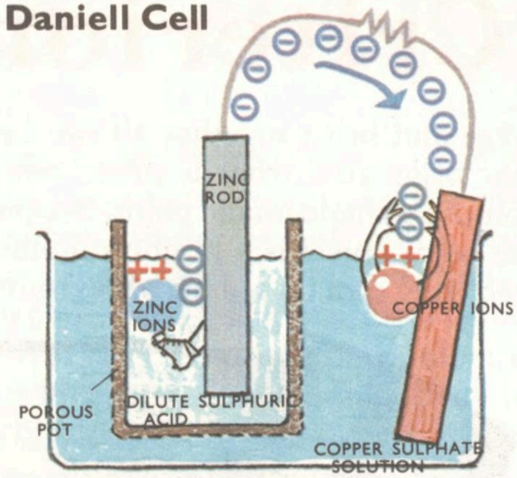
ANODE PELLET POTASSIUM HYDROXIDE CONTAINED IN ABSORBENT MATERIAL MERCURY-ZINC PELLET 2107

Polarization



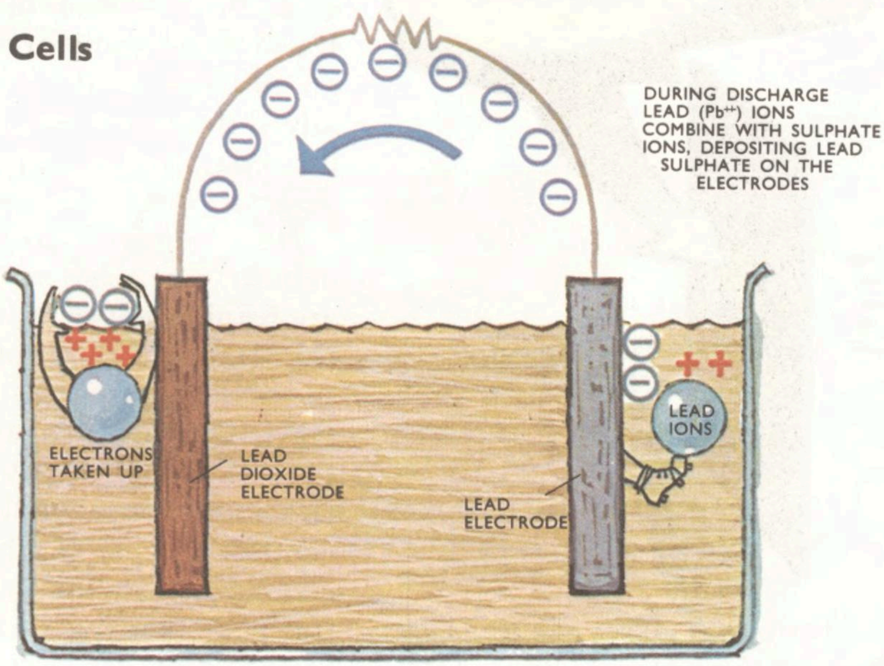
Any electrode that attracts positive charge will attract hydrogen ions which form bubbles of hydrogen gas, clouding the electrode and preventing the cell from working. To prevent this *polarization*, the cell must have some means of chemically removing the hydrogen ions before they have a chance to form polarizing gas bubbles.

Daniell Cell



The Daniell Cell can maintain a small current, delivered at 1·1 volts, for a long period of time.

Secondary Cells

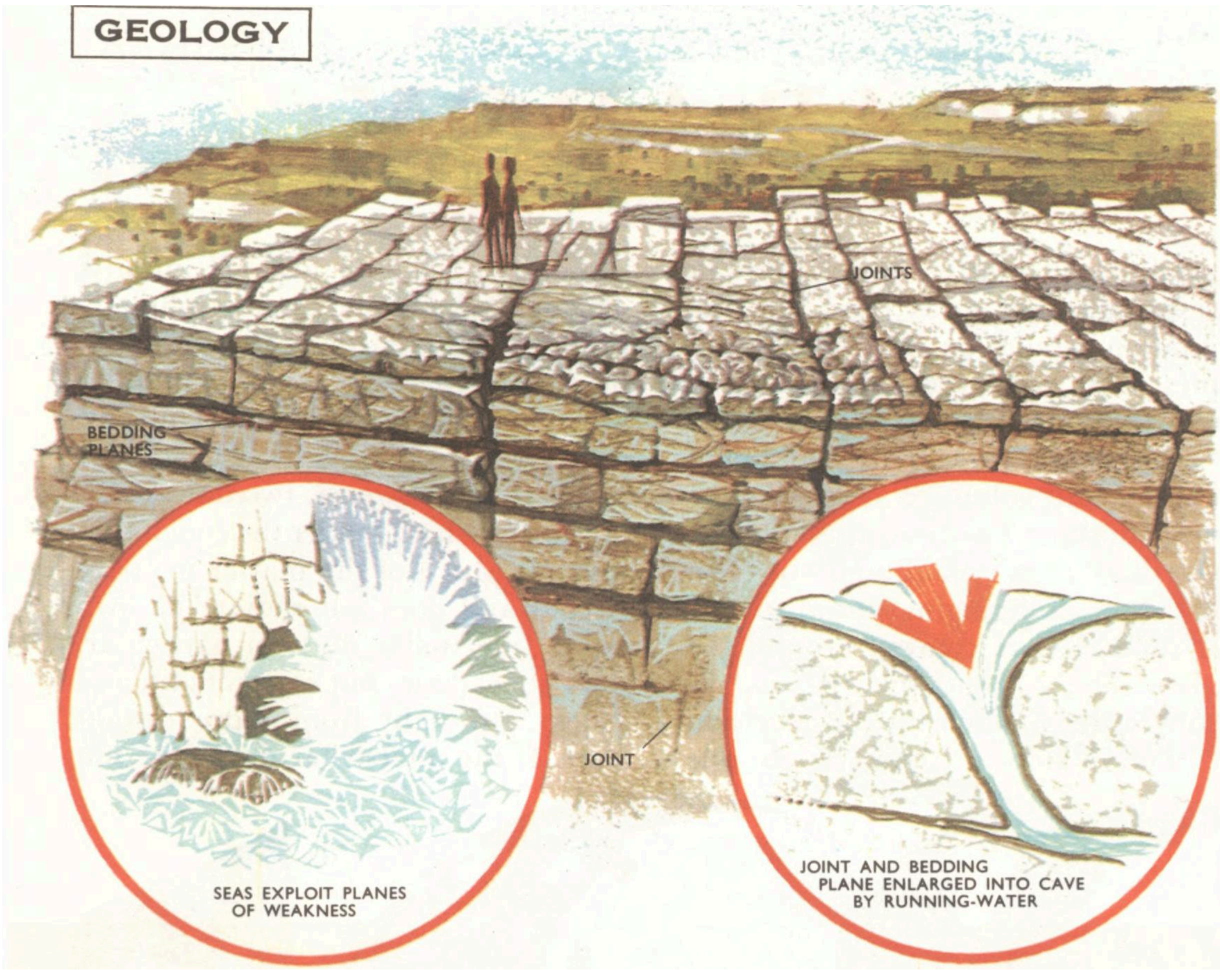


This cell, the lead accumulator, used for car batteries, discharges in a similar fashion to the primary cells. But the difference is that the chemical reactions taking place are *reversible*. Under discharge, the electrodes gradually turn to lead sulphate. When electrons are fed in the opposite way during charging the reverse reactions take place. The sulphate ions removed from the acid are put back and the electrodes become lead and lead dioxide once more.

Rayleigh's T is impossible to design and make a perfect optical instrument. The optical components — lenses and mirrors — invariably have defects and the designer has to bear this in mind and try to create an instrument in which the effect of these is as small as possible. The effects of the most important shortcomings in _lenses (chromatic and spherical aberration) have been largely eliminated in more refined pieces of optical equipment. But even when all possible care has been taken in designing an expensive Criterion verge and bring together all the light that originated from a single object point at a single image point. Suppose the point object is a brightly illuminated, tiny spot on a dark background. IN BRIGHT _ REGIONS LIGHT IMAGE OF SPOT HAS ALTERNATE LIGHT AND DARK FRINGES An image of the single bright spot is formed by the lens. Diffraction effects give the image a set of circular alternate light and dark rings around the central bright spot. (Inset) Light falls on each part of the image from all directions. In a bright region the light waves reinforce each other, in dark regions they cancel each other out. instrument a fundamental defect remains. This results from the wave properties of light itself, and it occurs as a result of diffraction — the break up of light wave fronts that creates interference patterns. This has an effect on resolving power. For example, if two small bright spots are being observed through a microscope, how far apart must they be before they appear as two separate spots and not as a single blurred image? Assuming that there is no aberration or other optical defect in the system, the degree of resolution is finally decided by the diffraction effects in the images of the two spots. An image is formed because light, which is given out in all directions, is collected by the lenses of the instrument. The lenses make the light con2108 This is giving out trains of light waves in all directions, each light 'ray' consisting of a series of 'crests' and 'troughs'. Some of the waves will be collected by, say, the central region of the lens, others by the outer regions. The lens brings together all of these rays to form the spot image. To form a replica of the small spot image, the waves passing through all the different parts of the lens must arrive in step — crest reinforcing crest, trough reinforcing trough. Away from the image spot there should be no light at all. Any crests that arrive there should be cancelled out by troughs. This is the theoretical ideal, but it never happens in practice. Not all the light waves that arrive in the image plane outside the central spot are cancelled out by other out-of-step In the design and manufacture of optical instruments there are a number of defects that have to be eliminated. By careful design, two of the most important, the aberration effects, can be much reduced. Chromatic aberration is caused by the fact that a lens focuses, at different points, light of different colours. An image with coloured fringes results from this. Spherical aberration occurs because light received by the outer regions of a lens is focused at a different distance from the lens from light that passes through near the centre. waves. Diffraction patterns result with alternate dark and light rings formed outside the central image spot. The existence of these outer rings spoils the resolution of an instrument. If two object points are viewed through an instrument they will not be resolved if they are too close together. If the separation is increased the image separates into two image points. It is desirable to be able to say definitely at what separation objects are being resolved by an instrument — this is a measure of the resolving power of the instrument. Lord Rayleigh suggested an arbitrary criterion for this. In the Rayleigh criterion two points are said to be resolved if the centre of the bright central disc coincides with the dark ring around the neighbouring image spot. If the image points are closer than this, they are unresolved. HERE Sao anor PEAK OF SECOND SPOT COINCIDES WITH TROUGH OF FIRST. POI ARE TW VERGE TO FORM THIS PATTERN Left, The brightness of the pattern can be plotted ona graph. There is a region of maximum brightness in the central circle followed by minima and maxima in the outer dark and bright rings. Right, According to Rayleigh's criteron the two points are resolved when the central bright maximum of one image coincides with first minimum of the L other.

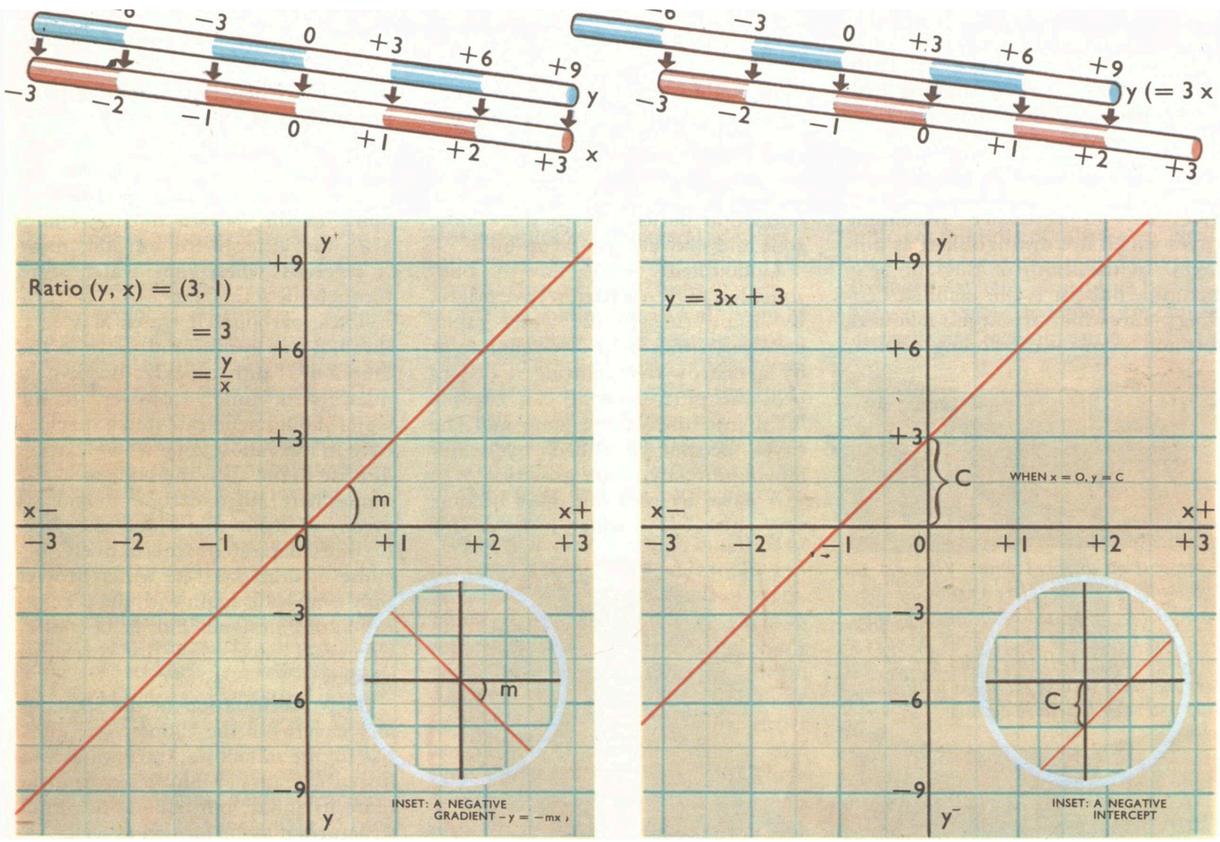
ASTRONOMY OLBERS' PARADOX S the Universe infinite, or does it end somewhere, with completely empty space and nothingness beyond ? Heinrich Olbers (1758-1840), a German astronomer, thought he had found an answer to the question. He argued that if the nearby parts of the Universe were a typical sample, repeated again and again in more distant parts, then it would be impossible for the Universe to be infinitely large. The Universe would contain an infinitely large number of stars and galaxies. Each star radiates light, and no matter how far away the star is, some of its light reaches the Earth. The argument continues: if the Earth were receiving light from all directions from an infinite number of stars, then the night sky would appear blindingly bright. In every direction there was bound to be a star. The sky would appear to be an unbroken source of light. In reality, the stars make a very small contribution to the light received on Earth. So according to Olbers, there cannot be an infinite number of them, and the Universe must be bounded. The more distant the star, the smaller the proportion of its light which eventually reaches the Earth. The contribution to the illumination on Earth falls off as the square of the star's distance. Quantities which fall off with squares of distance fall off very rapidly. If the Sun were twice as far away, the illumination would be reduced to one quarter. Three times its present distance away, and the illumination would be one ninth, four times away, and the illumination is down to a sixteenth of its present value. Over large distances, the number of stars in a unit volume is assumed to be fairly constant. Olbers considered the contribution to the illumination on Earth from an imaginary thin spherical shell of stars centred around the Earth. The entire Universe was divided in this manner into a succession of thin, spherical shells, each one enveloping all the smaller spheres. Olbers calculated the contribution from a typical spherical shell, and then totalled the contributions from all the shells — from the closest immediately surrounding the Earth, to the most distant. If the Universe were infinite, this outermost shell would be at infinity. But the paradox was that as the illumination decreased with distancesquared, so the shells became bigger, and contained more stars. In fact, the number of stars in each shell increased as the distance-squared. The increase cancelled out the decrease exactly, and left Olbers with a simple form of summing-up (or integration). If the distance of the most distant spherical shell were infinity, the result of the summing-up would also be infinity. The night sky should appear infinitely bright. Because it is dark, thesummingup cannot be taken as far as infinity. Olbers concluded that the Universe must be finite. There are, however, two flaws in his argument which render the conclusion invalid. Much of the matter in space is cold and does not radiate its own light. Interstellar dust, in particular, does not radiate, but absorbs light, and blocks out light from more distant stars. If the interstellar dust were not there, the Earth would receive much more light from the stars. The other flaw came with the discovery that all distant galaxies were receding away from our own Galaxy (this fact is shown by the red shift in their spectra). The more distant they are, the faster their speed or recession ; some observed galaxies are receding at half the speed of light (i.e. at go,000 miles per second). It seems likely that there are more distant galaxies, travelling at about the speed of light. Astronomers on Earth will never be able to detect these galaxies, because their light wall never reach the Earth. The galaxies could stretch out to infinity. None of the light from the most distant ones could ever contribute to the brightness of the night sky. 2109

SEAS EXPLOIT PLANES OF WEAKNESS JOINT AND BEDDING PLANE ENLARGED INTO CAVE BY RUNNING-WATER Limestone scarp showing bedding planes and joints. Such planes of weakness are exploited by weathering and caves may form. Inset — the sea and running water are the two important cave-forming agencies. ATURALLY-OCCURRING underground cavities are called caves — in distinction to tunnels and shafts which are man-made. Far more than just being intriguing 'holes in the ground', caves have brought a whole new field of science into existence — the science of speliology (Greek, spelaion, a Cave). Speliology embraces many scientific fields — geology, chemistry, physics and biology. The speliologist wants to know what the physical conditions are like underground, what organisms live there, what creatures formerly used the caves as shelters. But perhaps most fundamental of all, he wants to know just how caves have been formed. The Formation of Caves Every sort of rock contains openings. This property is called its porosity. The openings may be pore spaces between particles of sediment (primary porosity) ; or they may be fissures opened along planes of weakness (secondary porosity). The three planes of weakness important in cave formation are the bedding planes—lines of contact separating rocks of different composition and physical properties; faults — fractures in the crust along which slabs of rock 2110 have moved; and joints — fractures which have simply opened in rocks under stress. It is the planes of weakness rather than the pore spaces which are important in cave formation. The openings are initially very small. But they become enlarged by natural agencies — either the sea or running water. Formation of Sea Caves The sea makes caves in all types of rock. Pounding waves beating against cliffs exploit the planes of weakness. The water rushing against the rocks traps air in the small openings already present. The air, enormously compressed, adds to the battering action of the waves and the rocks begin to disintegrate. Any type of rock which outcrops at the coast is subjected to this treatment and caves can develop in all of them. However, the 'weak' rocks such as soft clays do not possess the necessary rigid strength and caves which begin to develop soon collapse. Larger, more enduring caves are cut into the harder rocks — sandstones, limestones, and igneous rocks — particularly where wave action is strong. A rise in the level of the shore may How Caves mean that former shore-line caves are lifted high above their places of origin. Formation of Inland Caves The action of moving water through the rocks is the chief factor in the making of inland caves. It is the corrosive action of the water — its ability to dissolve the rocks it is passing through — which is of greatest importance. The mechanical erosion by the moving water does not have the same marked effects. Consequently, caves are rarely found in igneous rocks or sandstones. They are nearly all confined to limestone country. Limestones are composed of at least 50% calcium carbonate, sometimes with a certain quantity of magnesium carbonate also present. Pure water does not have much effect on limestone. But in falling through the air, and especially in penetrating through surface layers of soil, rainwater becomes charged with carbon dioxide — it becomes a weak solution of carbonic acid. The carbonic acid reacts with the insoluble calcium carbonate, forming the soluble bicarbonate salt which is carried away. Even so, not all limestones necessarily make good caving country. The limestone must be thick and receive a sufficient quantity of rain water. Next, the planes of weakness must be suitably arranged. Where joints and bedding planes are close together, too many openings are available and no single The vadose zone lies above the water table, the phreatic zone below. In the vadose zone movement is controlled by gravity. In the phreatic zone movement 1s caused by water pressure and channels are more horizontal and may even slope uphill.


GEOLOGY
JOINTS
BEDDING PLANES
JOINT
SEAS EXPLOIT PLANES OF WEAKNESS
JOINT AND BEDDING PLANE ENLARGED INTO CAVE BY RUNNING-WATER

Are Formed large passage develops. What is needed are well spaced planes with flow-off locally concentrated by favourable ground contours. By corrosion, funnel shaped cave entrances may develop — the swallow holes or sink holes. In Britain the thick Carboniferous limestones are most suitable for cave formation. Chalk, a very pure limestone, is too 'weak' and caves soon collapse. Important in predicting just where caves are likely to develop in a limestone area is consideration of the water-table. Below the table, all cavities in the rock are permanently filled with water. This zone is called the phreatic zone. Above, rocks are not saturated —this is the vadose zone. Water simply passes through into the phreatic zone below. The obvious place to expect caves to originate and develop is in the vadose zone. Water, charged with carbon dioxide, moves downwards under the influence of gravity. Paths followed may be along dipping bedding planes, or they may be down almost vertical joints. In the north of England, vertical joints have commonly been followed by water. This gives rise to vertically descending caves which are distinguished as potholes. In the south of England, e.g. Mendip hills, it is the inclined bedding planes which are largely followed, forming downward-sloping — rather Cave systems are not static. Changes are constantly occurring. Here a newly developed channel has by-passed an original passage, with the result that the original passage is no longer used. Left 'high and dry' it may over the years become encrusted with stalactites and other limestone deposits. than vertical descents. In the phreatic zone water movement also takes place, though not nearly so fast as in the vadose zone. Water is finally discharged by seepage into seas or rivers. The movement is caused not so much by gravity as by hydrostatic pressure — the water head. The pressure arises because the water table at the points of outlet are at a lower level than elsewhere in the table. called a cavern. Heavy falls of rock may, however, appreciably block an exit and the water becomes forced to flow out through another passage. The initial pattern may as a result become altered. Some roof collapses give rise to depressions in the ground above — the shakeholes. Extremely large scale collapses may actually form gorges. Structures in Caves Pure water has little effect on limeA CAVE IS ENLARGED CAVERN Caves become enlarged by seepage and roof-fall into caverns (right). Complete roof collapse along a line of caverns may result in a dry limestone gorge. Left, the vertical-sided Cheddar Gorge, Mendip Hills, Somerset may well have been formed by just such a sertes of collapses. Flow against gravity may even occur for water may at times be moving at a lower level than the ultimate point of discharge. Caves originating in the phreatic zone generally follow more horizontal paths than in the vadose zone and may even slope up-hill. Undoubtedly water flow in both zones has played a part in cave origin. By a lowering of the water table, however, either by a drier climate or by a nearby river cutting a deeper bed, the phreatic zone can become lifted into the vadose zone and the caves become modified by water moving directly under gravity. A cave, by rock fall, may become very enlarged, in which instance it is stone. Only between 20 and 74 parts per million of calcium carbonate by weight will be dissolved. In contrast, water containing carbon dioxide gas dissolves up to 400 parts per million. Some of the carbon dioxide comes from the atmosphere but far more is received when rain water seeps through the soil. Thus, carbonated water may seep through the fissures and pores in the limestone, and quickly dissolve a quantity of lime. However, if the water should penetrate into a spacious cave in the vadose zone where carbon dioxide in the air is the same as in the atmosphere, then carbon dioxide is given off from the water and lime becomes deposited. Alternatively, because of draughts, the water droplet may evaporate again with the precipitation of limestone. This is, of course, the exact reverse of the corroding process. Such precipitation in the spacious airy caverns of the vadose zone gives rise to the stalactites, stalagmites and frozen cascades. Many lesser known, more irregular, but equally beautiful structures may be formed as well. 2111 BY OVERHEAD

MATHEMATICS GRAPHS and EQUATIONS ‘THE graph drawn from experimental results may turn out to be a smooth curve, or even a straight line. From the shape of the graph, the steepness of its slope, the places where it crosses the ‘x’ and ‘y’ axes it is possible to analyze the results of the experiment, and perhaps deduce the law holding for it. A graph shows how two quantities vary at the same time. Straight lines show that one of them is proportional to the other. Steps of equal size in the ‘x’ direction correspond to steps of equal size in the ‘y’ direction. Another way of saying that they are proportional is that the ratio of the set of ‘y’ values to the set of ‘x’ values (corresponding members of sets) is constant throughout. Often, on a graph, the point where the two axes cross (called the origin) is at x = 0, y =o. If the graph is Two parallel number lines can show when two sets of numbers are proportional to each other. The conventional graph consists of two number lines, an ‘x’ line and a ‘y’ line, at right angles. = ae ~ | LT y] > Proportional sets of integers. On the graph, the correspondence between the points is shown by a straight line through the origin. a straight line, and it goes through the origin, then ‘x’ and ‘y’ are related by a very simple equation _y = mx. The sets of values are proportional. The ratio is constant. y:x is the same for all corresponding values of y and x. This ratio value, a number which remains constant throughout the sets, is called m. .y:x = y +x =m. To get to y = mx, both sides are multiplied by x. Often straight lines do not score a direct hit on the origin. It misses it and crosses both the y and the x axes. Even then, it is possible to deduce the equation of the line. For example, where the line crosses the vertical _y axis The ‘y’ line is slid backwards. Now o on the ‘x’ line is alongside 30n the ‘x’ line. y = 3x + 3 for each pair of corresponding points. y (= 3x +3)



y
x
y (= 3 x
Ratio (y, x) = (3, 1)
= 3
= y/x
m
INSET: A NEGATIVE GRADIENT – y = –mx
y = 3x + 3
C
WHEN x = O, y = C
INSET: A NEGATIVE INTERCEPT

is important. The y axis marks all points where x = o. When x = 0, y has some value. y must be something since it was decided that the line would not go straight through x = 0, y = 0. This value is called the intercept and is given the symbol ¢ . It stands for a number. The equation for the line has therefore an additional bit in it. It becomes y = mx + c. When x = 0, mx = 0 and y = c. This does happen on the graph. Curved lines present more problems. It is difficult to. get some definite equation out of them. However, curved lines can be used in the opposite direction, going from an equation to a graph to solve the equation. An equation which contains numbers x’, x”, x*, and so on, can be easily drawn on a graph. It is probably much more difficult to solve by algebra. For example, x? — x — 2 = o. What is x? Instead of solving the equation by algebraic means, x? — x — 2 is made equal to ‘y’, the quantity up the vertical axis of the graph. So y = x — x — 2. Then x can be given any reasonable values we wish f= 2 1 O Fl weeietgs a4 P= t+4 +1 Q > Agito —Kee Tas at GO. <1 wht eeg ne Baw) fh O:. 228.08 a) 4 Graphs and Proportion The sets of numbers -—-9 -6 -3 0 +3 +6 -—-3 —2 -I 0 +1 +2 are proportional and lead to a straight line graph. The sets of numbers +9 +4 +1 ee ee 0 +1 +4 +9 Cc +1 +2 43 are not proportional and lead to a curved graph instead of a straight-line graph. Numbers in the upper set are all the squares (i.e. the number multiplied by itself) of the numbers in the lower set. Enough values are taken for a smooth curve to be drawn. The curve crosses the x-axis in two places. These are the solutions of the equation. Why should the solution be given by the points where the curve crosses the x-axis? The reason is that for at all points along the x-axis, y = o. Our graph shows y = x? — x — 2. But we are especially interested in 0 = x? — x — Q, i.e. the values of x when J) = 0. If a curve of this form cuts a line at all, it cuts it in two places (except when it just touches the curve — this is a special case). It could happen that the curve stays clear away from the x-axis. This is an indication that there is no real solution to the equation. If there are two cutting points, there are two solutions. This always happens when there are x? terms in the equation. If it includes x? terms, then there would turn out to be three points, with three solutions to the equation. “ a “< rnynwrkun © 0 Altering the scale of either the ‘x’ or the ‘y’ number line has no effect on the ratio, m, of the proportional sets. Graph scales:are chosen so that the graph fits onto the page. BEBE SESE ee VET ZT TU sébeidn | | BEN EE? eRe wes Be SS 42a

RREs Using a graph to solve the equation 0 = x? —x —2, a quadratic equation which gives two possible values for x. The solutions are found from the points where the curve y = x' —x —2 cuts the ‘x’ axis of the graph. Left: when a quadratic curve touches the ‘x’ axis, the two ossible solutions coincide. Centre: a single solution, x = 0. ight: the curve shows that the function represented by ‘y can never be equal to 0. The equation has no solution.

Graphs and Proportion

The sets of numbers

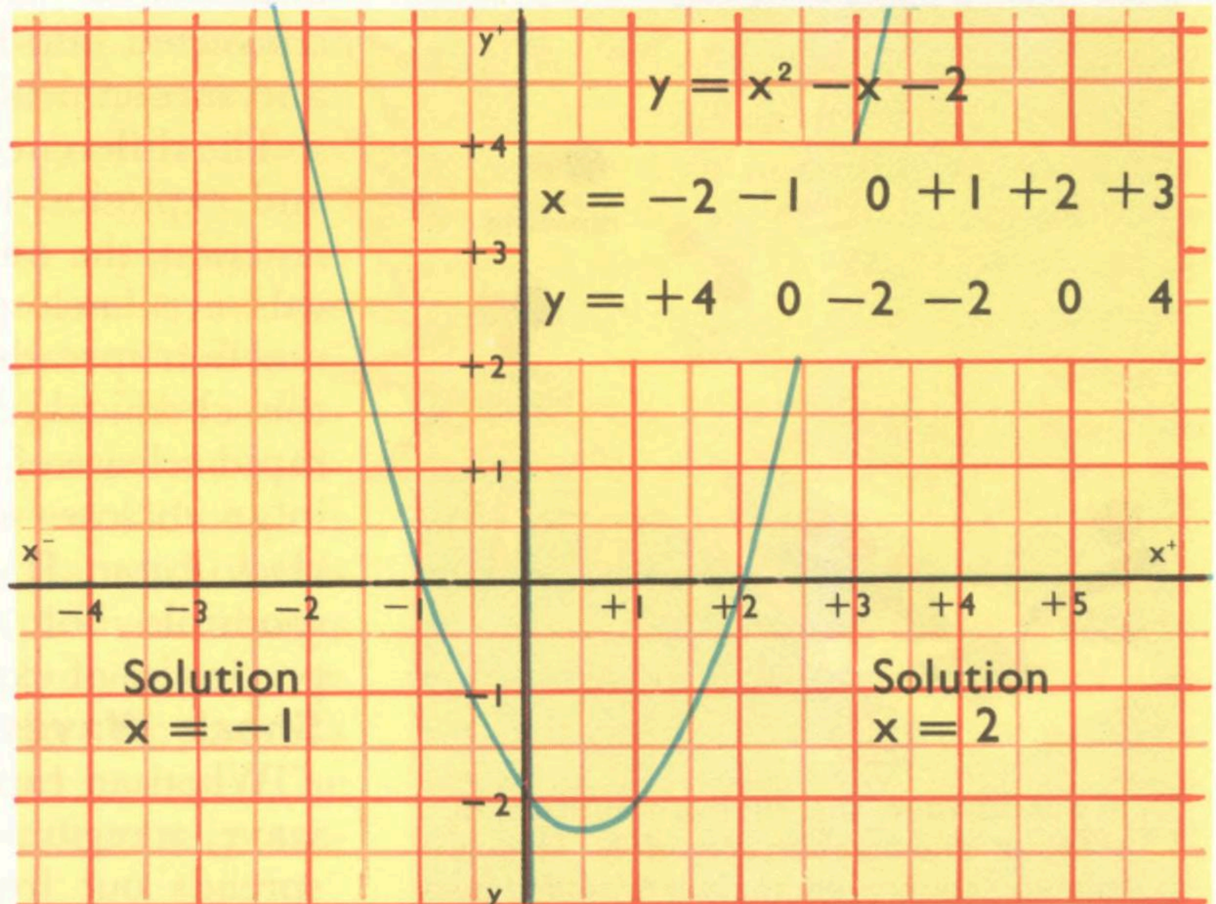
−9	−6	−3	0	+3	+6	+9
−3	−2	−1	0	+1	+2	+3

are proportional and lead to a straight line graph.

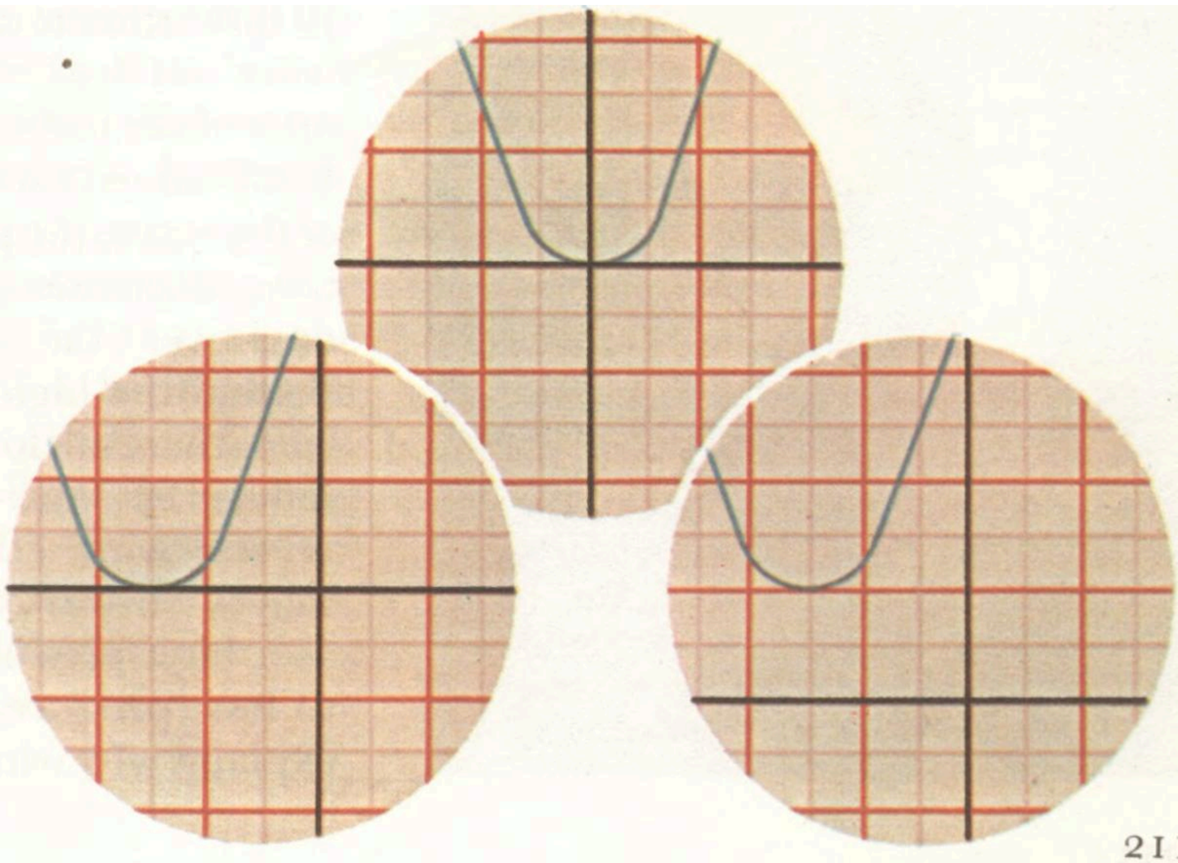
The sets of numbers

+9	+4	+1	0	+1	+4	+9
−3	−2	−1	0	+1	+2	+3

are not proportional and lead to a curved graph instead of a straight-line graph. Numbers in the upper set are all the squares (i.e. the number multiplied by itself) of the numbers in the lower set.



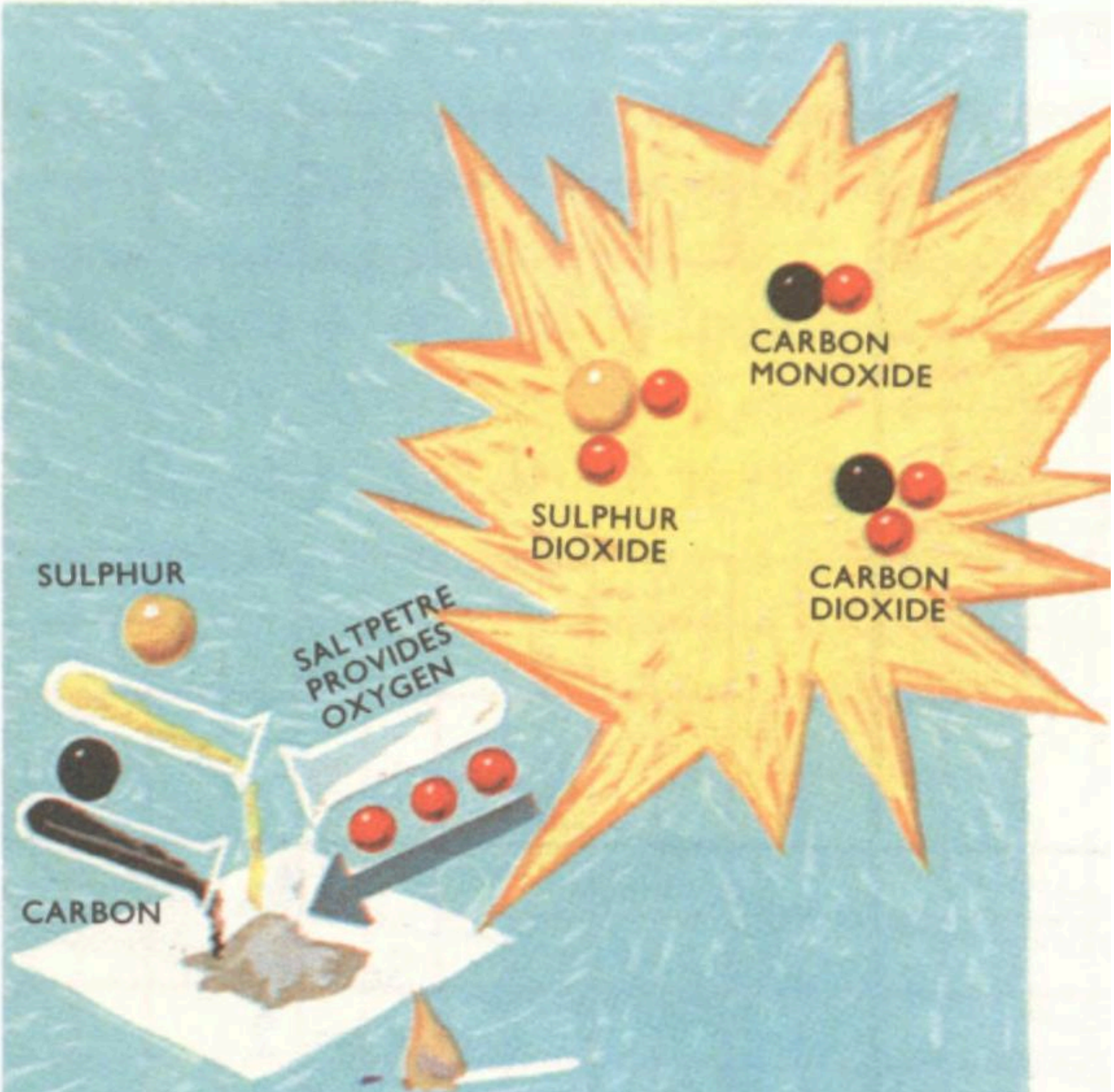
Using a graph to solve the equation $\mathbf{0 = x^2 - x - 2}$, a *quadratic* equation which gives two possible values for **x.** The solutions are found from the points where the curve $\mathbf{y = x^2 - x - 2}$ cuts the '*x*' axis of the graph.



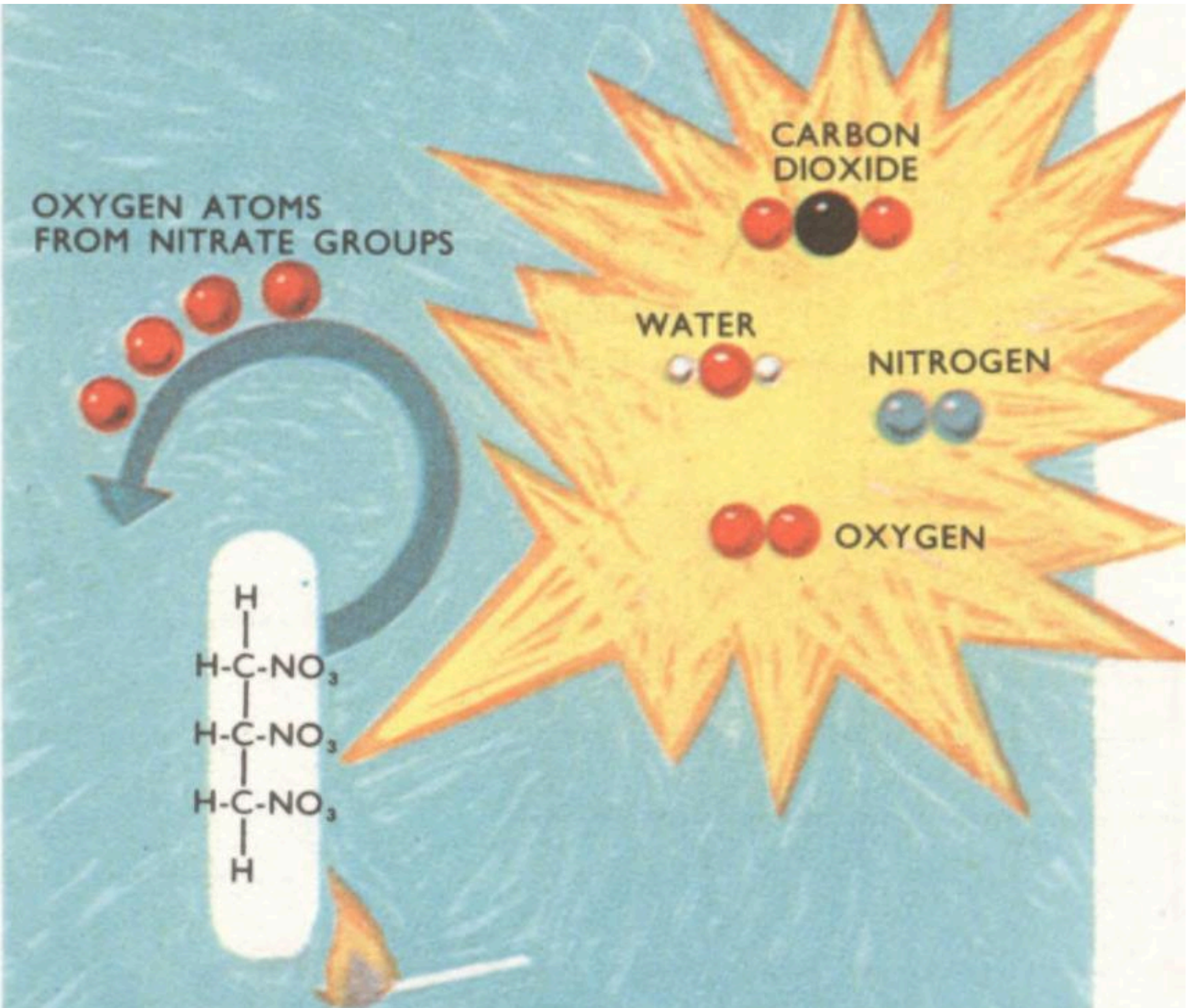
2113

| APPLIED SCIENCE | EXPLOSIONS HE science of explosives has advanced a long way since the day of the primitive cannon. Then, crude cannon balls were hurled at the enemy, projected by gun-powder, detonated by lighting fuses. In the twentieth century, as a result of the combined work of many of the greatest living scientists, nuclear explosives have been developed that can shift mountains and divert the flow of 2114 rivers, or can blast out of existence whole islands or cities in a fraction of a second. In a nuclear explosive, the vast amount of energy locked away in the nuclei of atoms is suddenly released. But what happens when a chemical explosive is detonated ? It is not very different, in principle, from ordinary combustion or burning that occurs when a match is put to tinder. Chemical energy is stored in the explosive and this is suddenly released and converted into large amounts of heat and mechanical energy. The difference between combustion and explosion lies in the rate of the reaction, the way in which the chemials are enclosed, and the way the reaction spreads through the body of the chemical. In an explosion, the rapid release of large quantities of gas in an enclosed space sets up powerful shock waves. It is these that are responsible for all the well-known properties of explosives. Shock Waves When an explosion occurs, a shock wave, a region of very high pressure, spreads out from the source. Shock waves travel at least as fast as the speed of sound — the higher the pressure in the wave, the greater the speed. In air, the energy of the explosion is dissipated not only in the movement of air (wind) caused by pressure difference between the shock wave and its surroundings, but as heat, generated by the moving gas. The shock wave front spreads out in three dimensions from the explosion 'core', and as it does so, the surface area of the front increases, so the effect is spread over a large area. The effect of the wave, the pressure and temperature, decreases as the wave spreads out over the larger surface area. Behind the high pressure shock wave front there is a low pressure region. Air moves into this 'trough', causing a reverse wind effect. So, as the front moves outwards it is felt first of all as a high pressure 'push' followed by a low pressure 'suck' — the suction explains why windows are often sucked Shock waves may be studied in the laboratory in shock tubes. The shock wave 1s created by releasing a high pressure pocket of gas in the driver section of the tube, by breaking a retaining diaphragm (inset). The wave then proceeds down the tube. The progress of the wave is studied using

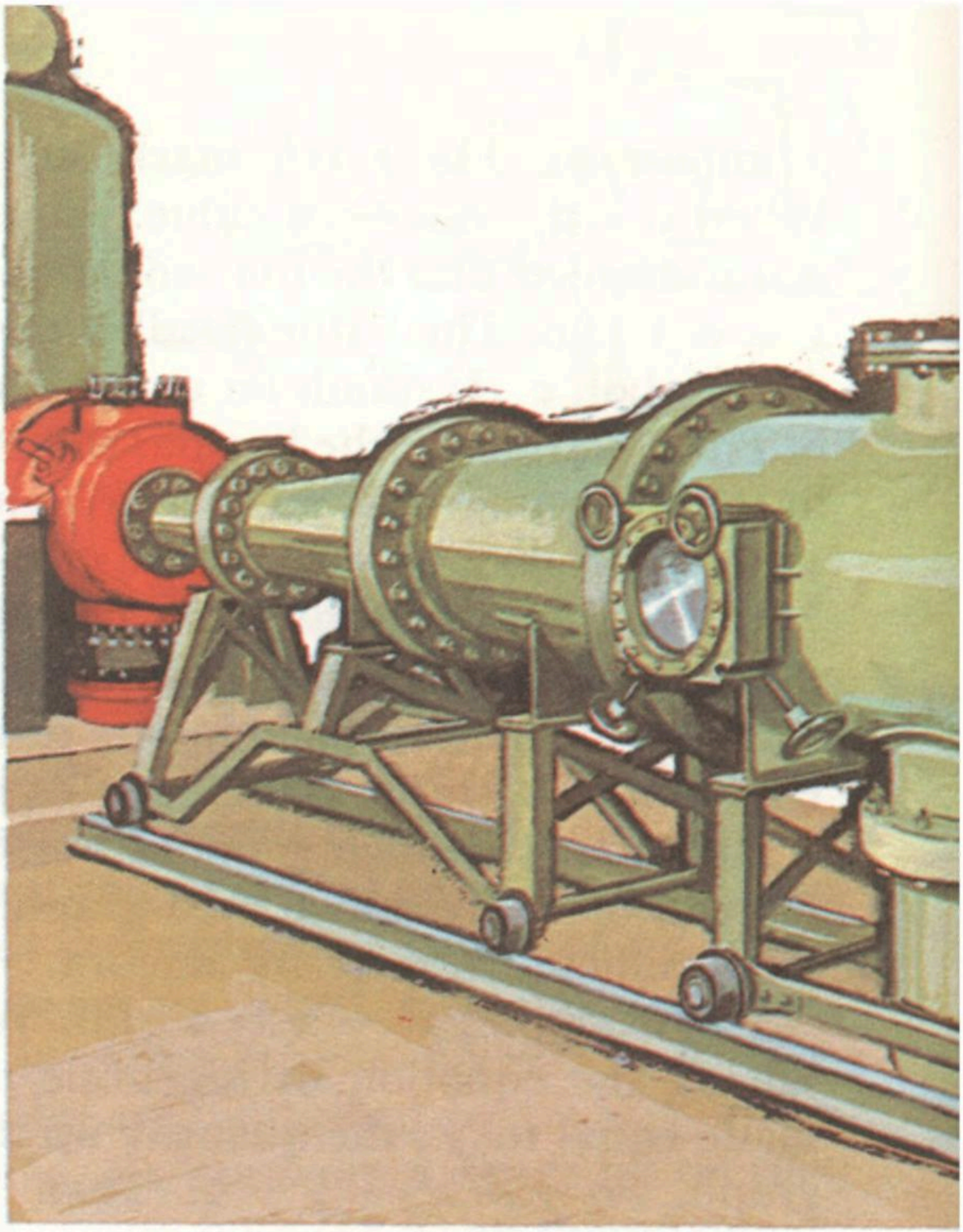
optical (interferometric) instruments. outwards as a result of an explosion in the neighbourhood. Similar shock waves are formed when explosions are set off under water. The effect of an explosion does not only depend on the amount of explosive used, although this is of prime importance. The degree of confinement of the explosion is also very important. A slow seepage of combustible gas in the open will perhaps produce a large sheet of flame and a After an explosion a shock wave spreads out Srom the ‘core’ with speeds greater than the speed of sound. The high pressure region is immediately followed by a low-pressure ‘suction’ region.



In gunpowder, the 'fuel' is a mixture of sulphur and carbon powder. This is 'burned' in oxygen, mainly provided by saltpetre (potassium nitrate). The resulting explosion is due to release of large quantities of sulphur dioxide, carbon dioxide, and carbon monoxide.

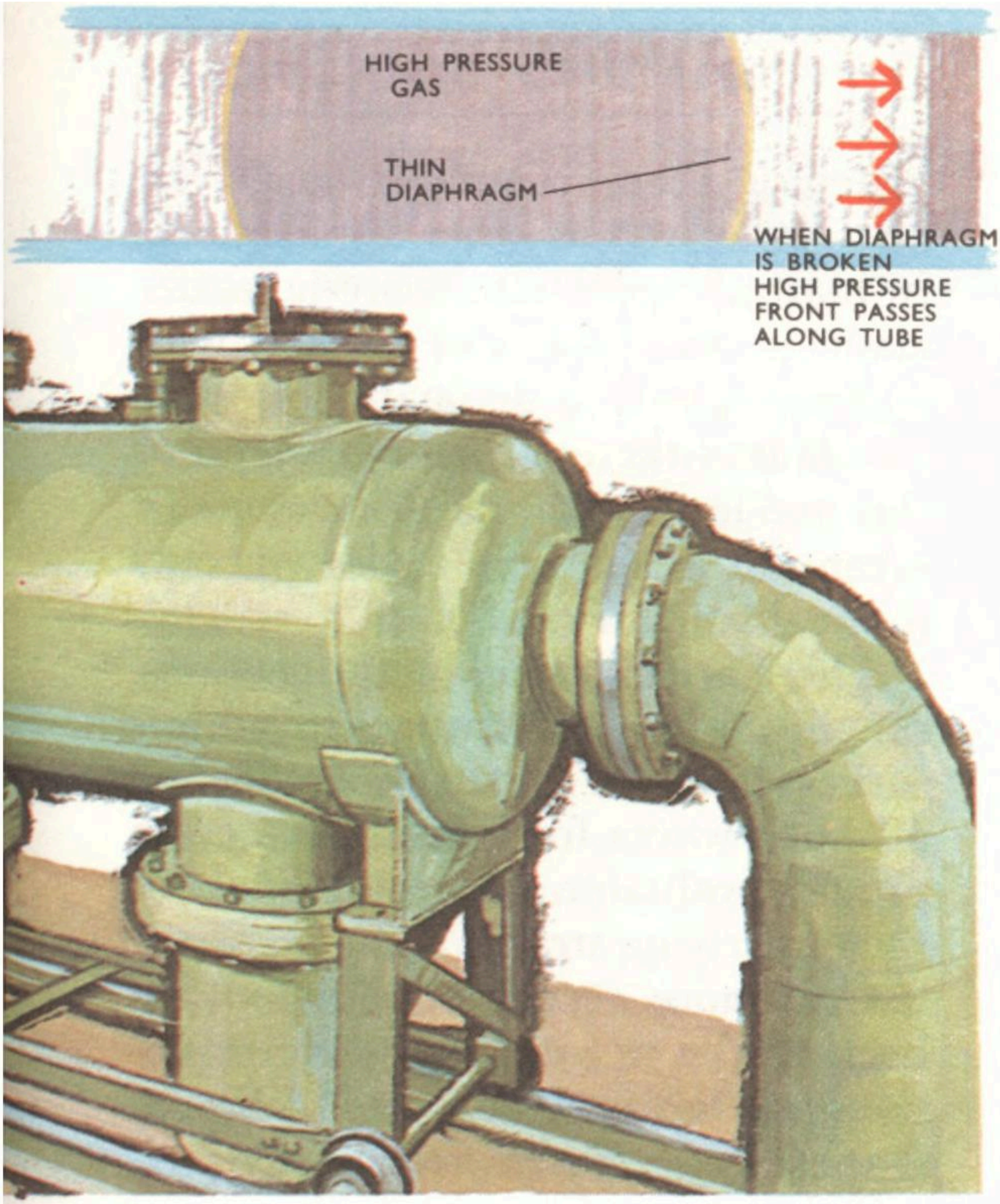


When a pure explosive like trinitroglycerine explodes, the oxygen is provided from within the molecule of the compound itself. The carbon hydrogen and nitrate groupings are the fuel. A mixture of gases – carbon dioxide, carbon monoxide, nitrogen, oxygen and water vapour are released.



HIGH PRESSURE FRONT PASSES ALONG TUBE noticeable updraught, but if it is confined as is firedamp (methane) in an underground mine, a disastrous explosion might result. Likewise, petrol vapour from the wick of a petrol lighter burns steadily when ignited by the spark from a flint. When a petrol-air mixture is spark-ignited in the cylinder of an internal combustion engine, it explodes. The difference here, once again, is that the petrol is being consumed in a confined space in the engine cylinder but not in the lighter. The principal difference between combustion and explosion is in the rate and mode of travel of the burning ‘front’ in the gases. In ordinary combustion, the flame front travels slowly, as the heating effect passes along the gas. Heat passes from one layer of gas to the next by thermal conduction. The flame front’s progress can be easily seen and measured, and flame propagation speeds of about 50 centimetres per second are normal. In a more confined space, the burning which causes gas to be formed builds up a high gas pressure behind the flame front. This pushes the front forward at a high speed. The gas becomes turbulent and movement within the gas makes it heat up still further. This all adds up to even greater acceleration of the flame front and a very high propagation speed of the flame front is reached, perhaps 3,000 times greater than in an ordinary open flame. Although the exact physical. mechanism that governs an explosion is not altogether understood, it is clear that the build-up of high pressure in the confined space is an essential part of it. The Chemistry of Explosives An explosive is essentially a very fast-burning ‘fuel’. In the explosion, oxygen is rapidly consumed and the fuel, a complicated chemical compound or mixture of compounds, is coverted into a mixture of simple gases like carbon monoxide (CO), carbon dioxide (COg), water vapour (H,O) or nitrogen (N2). Because each large (exploding) molecule is converted into a large number of smaller molecules there is a rapid increase in gas volume. Also, the reaction is strongly exothermic, so great amounts of heat are released. The efficiency of the explosion not only depends on the quality of the fuel but on the ready availability of large quantities of oxygen. The ordinary atmospheric oxygen is not sufficient to feed an explosion, although it can feed an ordinary flame. For this reason, chemical explosives always contain an oxygen-rich chemical — oxidizer. In gun-powder, the fuel is a mixture of sulphur and carbon. The oxygen is provided by saltpetre (potassium nitrate KNOs). In more modern explosives, ammonium nitrate

(NH4NQOs) is used as the oxidizer. This can be -used in conjunction with a large range of carbon-containing fuels (waste products of cereal-processing factories, for example). In some forms of dynamite, the ammonium nitrate is made much more effective by the addition of one of the pure explosives. In pure explosives, the fuel and the oxygen are provided in one single molecule. Compounds like _ nitroglycerine and trinitrotoluene (TNT) contain nitrate (NOs) groups that are capable of providing oxygen to feed the combustion of the rest of the molecule. Carbon and hydrogen in the molecules are oxidized to form a large number of small gas molecules — carbon dioxide, carbon monoxide, water vapour, nitrogen, and hydrogen. The explosion of nitroglycerine can, for example, be represented by the equation: Explosions may be caused in underground mines when firedamp (methane) collects in sufficiently high concentrations. It can be detected by the safety lamp, devized in its original form by Sir Humphry Davy. A more modern version 1s tllustrated above. 4C,H,N,O, = 12CO, + I0H,O+ 6N,+ O,, so it can be seen that 29 gas molecules are formed for every four original molecules in the solid. Consequently there is an enormous increase in volume caused by detonation of the nitroglycerine. Nitroglycerine is a very sensitive explosive and only became of practical use when Alfred Nobel discovered that it could be desensitized by absorption in a form of clay (Kieselguhr). T.N.T. is one of the most useful explosives, both in industrial and military applications. It is, however, comparatively short of oxygen even though it is a pure explosive and is made much more effective by the addition of d ary explosives, en by the rapid combustion fuel" ‘becomes available as a result of the re“arrangement of atoms in the original compounds to form different com4 pounds. In nuclear explosions, the atomic es are altered. In the ae - nuclei themsely 2115


HIGH PRESSURE
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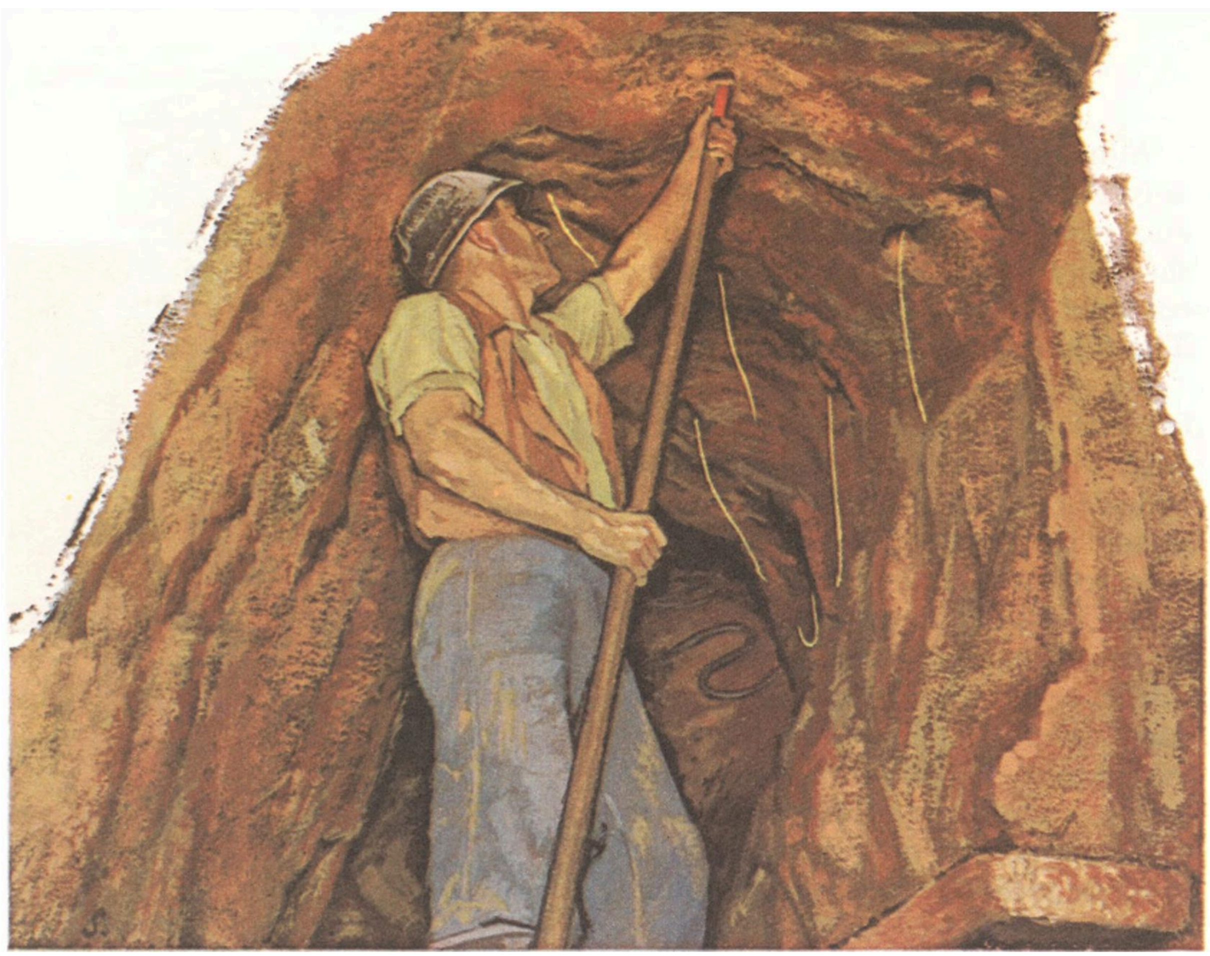
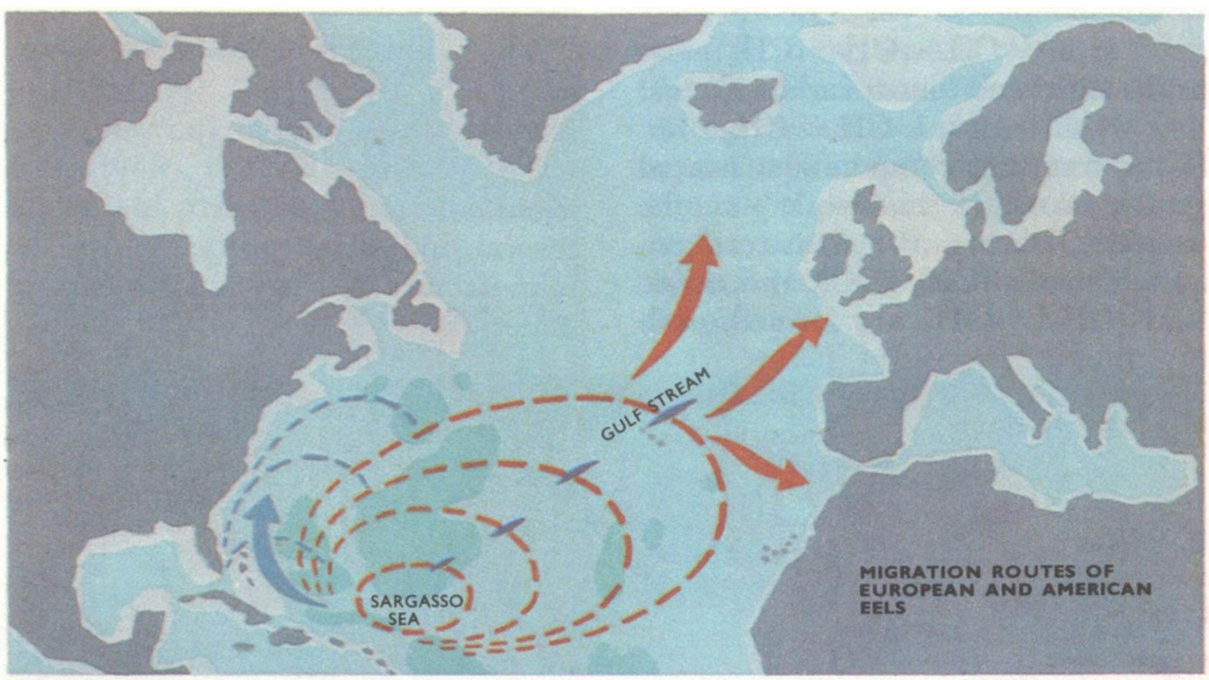
d Fo Ry Preparing for a blasting operation in an underground mine. Shot holes are charged with dynamite sticks. These are detonated by remote control. ammonium nitrate as an oxidizer. The resulting mixture is called amatol. T.N.T. has an advantage over other explosives in that it can be melted at fairly low temperatures into the shapes required to fill bombs and shells. It is also fairly insensitive to moderate mechanical shock and can therefore be handled with some degree of safety. Explosives such as T.N.T. and nitroglycerine are themselves detonated — set off— by powerful explosive shock waves that originate from primary explosives (detonators). The Explosives are used in seismic studies of riigalagen rock strata. The shock waves are reflected at the interface of the two strata and by studying the time taken for the shock pulses to reach the differently-placed microphones the position and nature of the underlying stratum can be found. » As & Ms MN REFLECTED PULSES INITIAL MICROPHONES ws 2116 detonator, a highly sensitive explosive material like lead azide (Pb(N3)q), or mercury fulminate (Hg(ONC),) is itself set off by mechanical impact or heat from a fuse wire. Propellants Propellants, used to eject bullets or rockets, are not explosives in the true sense of the word. The object is to produce, by steady but rapid burning, a volume of compressed gas that ejects the missile. The pressure build-up must occur in a very short space of time after the 'powder' is ignited. (The 'powder' is, in fact, often a solid shape in the form of a disc or cylinder.) The propellants used in large calibre guns are normally nitroglycerine compounds, in the form of cordite or ballistite. Cordite is basically a mixture of nitroglycerine and nitrocellulose with a sensitizer added. In the breach of a gun a gas pressure of 50,000 pounds per square inch might be generated in less than one five-hundredth of a second. In rocket projectiles, liquid propellants are often used, as well as solid ones. Liquid propellants are normally a mixture of fuel, such as kerosene, together with an oxidizer such as liquid oxygen or hydrogen peroxide. Explosives in Use Apart from the more obvious military applications, explosives are useful materials used in mining and rock blasting as well as other applications. In blasting operations, bore-holes are made in the rock and packed with explosives, with a detonator inserted. The hole is then packed with sand or clay and leads to the electric detonator taken out of the hole and connected to the detonatary switch. There are many other applications, ranging from steel cutting to seismic surveys in which the study of reflection of shock waves from inside the earth provide valuable information about the underlying strata. One more recent application has been in the use of small-scale explosions to form metals into different shapes. In explosive forming, as it is called, a sheet of metal is laid over the POLYTHENE BAG Explosive metal forming. The large forces generated in explosions can be used to form metals. In this arrangement, a charge is detonated in water contained in a polythene bag. The metal workpiece is formed into the shape of the die. open end of a die and is enclosed, together with an explosive charge, inside a sealed chamber. The highly compressed gas released by the explosion forces the metal into the die where it is formed into the required shape. This method is particularly useful in forming complicated shapes. It can also be used to cold-weld dissimilar metals together and to form solid sintered metals out of powdered metal.



INORGANIC CHEMISTRY OZONE XYGEN is used by all forms of living creatures. It is by far the most abundant of the elements, and approximately one fifth of the atmosphere of the Earth is made up of oxygen. When we breathe it in, it is used to convert food into energy. The oxygen in the air is normally in the form of diatomic (Og) molecules — twin pairs of oxygen atoms. But another form (allotrope) of oxygen can exist and this is in the form of triatomic molecules (O3) called ozone. This has very different chemical properties from ordinary oxygen and is quite poisonous. It is found in the atmosphere in very small quantities — the highest concentration being at high altitudes where it is formed by the action of ultra violet light on oxygen. Chemical Properties of Ozone The ozone molecule is a high-energy version of ordinary oxygen. Oxygen molecules have to be energized (usually electrically) before they come together to form ozone: 30,2 + energy = 2QO3. When the ozone molecule is formed it is active and unstable —it easily reverts to ordinary oxygen. It is a very powerful oxidizing agent. For example, lead sulphide is oxidized to lead sulphate by ozone: PbS + 403 = PbSO, + 402 and ferrous sulphate (in the presence of dilute sulphuric acid) to ferric sulphate: ; 2FeSO, a Os 5 H,SO, = Fe2(SQ,4)3 + HzO + O2 In these reactions, the ozone molecules lose a solitary oxygen atom, but in other reactions the whole molecule takes part. For example, sulphur dioxide is oxidized to sulphur trioxide, using all three oxygen atoms in the process : OXYGEN IN Producing ozone in laboratory. Oxygen is passed into the ozone tube. This consists of two glass tubes, the outer surface of the outside tube and the inner surface of the inside, each coated with tin foil. High voltage sparks are passed between the glass surface through the oxygen, to form ozone. 350, + Os = 3803 Although ozone is present in very small proportions in the air its powerful oxidizing action can, over a period of time, cause cracking in rubber tyres — to overcome this, anti-oxidants are mixed with the rubber. Ozone Finds the Double Bond Important classes of organic compounds contain double or triple unsaturated chemical bonds, and the chemist needs to know where these are situated in the molecule. Ozone is used in analysis because it is taken in at the unsaturated link to form explosively unstable ozonides. The compound under investigation is dissolved in an organic solvent and ozone is bubbled through to form the ozonide. In butylene, for example, the ozonide link can be formed either between the second and third carbon atoms at the double bond. H,C — CH = CH — CH; or the third and fourth carbon atoms: H,C ae CH, ca CH => CH, If the resulting compound is heated gently in water it 'breaks' at the ozonide link. In the first reaction, acetaldehyde (CH3;CHO) and acetic acid (CH;COOH) are formed, each with two carbon atoms. In the second reaction two compounds, one with three carbon atoms and one with a single carbon atom, are formed. By identifying the resulting fractions the position of double bond in the original compound is found. Preparation and Physical Properties of Ozone To form ozone, ordinary oxygen is passed through a tube where it is subjected to silent electrical discharges. The electrodes are covered with insulating materials (glass or mica) so that the discharge currents are kept low, and high voltage pulses are applied between the electrodes. This method has to be used so that the oxygen acquires the energy it needs to form the ozone, but at the same time the ozone does not become so hot that the molecules break up into ordinary oxygen. The gas itself is light blue in colour, and has a strong odour. It is used to sterilize water, and to prevent bacterial growth in food stores. Its ozonide-forming property is used in several industrial processes, in the manufacture of drugs and plasticizing materials. Ozone as an Oxidizing Agent Finding the Double Bond Lead sulphide is oxidized to lead sulphate by ozone — a very powerful oxidizing agent. H H eae H re) fe) eS ee > / Ss ee 4 4 ee 1: HO, Sd gue H—C +H-C C} LEAD LEAD SULPHATE eo H ozone 4 H H OH @®@ wh me is — esi ¢ c "re ethylene Seoiige formaldehyde formic acid o. ee OxYGEN Ozone breaks into double bonds in organic compounds, forming ozonides. The ozonide is unstable and is easily split into two compounds. Identification of the two compounds formed indicates the position of the double bond in the original compound.

BIOLOGY The INCREDIBLE EEL [N 1856 Dr. Kamp, a naturalist, captured an unusual salt-water fish. Somewhat like a delicate, transparent laurel leaf in appearance, the fish resembled no fish previously known. Dr. Kamp named it Leptocephalus brevirostris. Nearly forty years later, in 1895, two Italian naturalists, Grossi and Calandruccio, decided to find outa little more about Dr. Kamp's new species. Several leptocephali were captured in the Mediterranean near Messina and placed in an aquarium. They readily ate the food given to them yet, instead of growing in size, they became smaller. Their transparent, leaflike appearance was gradually replaced by that of a thinner, more 'opaque creature until it became obvious that Dr. Kamp's leptocephali had changed into elvers or 'glass-eels'. Leptocephalus was not therefore a valid genus, it was merely the larval stage of the eel, Anguilla. Grossi and Calandruccio reasoned that the eels, upon reaching maturity, returned to the sea to lay their eggs at the bottom: and, since no adult returning to rivers had been captured, concluded that the adult eels died at sea. The young eels stayed at the bottom until they were almost ready to change into the elver stage. As elvers they swam upstream into the rivers. Grossi and Calandruccio explained the rarity of the leptocephali by stating that few were caught in the fishermen's nets because they stayed near the bottom. Specimens could be obtained from the straits of Messina because here current upwellings brought deep water creatures to near the sea surface. But the mystery of the eel's life history was far from being solved. Certainly Grossi and Calandruccio's work had once and for all destroyed previous myths about the origin of eels from such sources as 'the hairs from horse tails', 'grass sod' and other unlikely materials. There were such things as young eels and it was probable that the eels merelydid the reverse of such fish as salmon and sturgeon, namely returning to the sea to spawn, but spending their life as adults in fresh water streams and rivers. In 1904 a Danish biologist, Dr. Johannes Schmidt, was on board a Royal Ministry of Fisheries Steamer, the 'Thor', fishing between Iceland SS and the Faroe Islands when a transparent leptocephalus was caught by a surface net. Its length, about 3 inches, was comparable with that of the specimens captured by Grossi and Calandruccio at Messina. Off the Irish coast, a few months later, one more leptocephalus was captured, again at the surface. It was certainly not true therefore that all leptocephali were restricted to deep water. As a result of his finds Dr. Schmidt was commissioned to investigate further the eel's mysterious life cycle. He found that the 500 fathom line on nautical charts — this marks the edge of the continental shelf — roughly coincided with the areas where threeinch leptocephali had been captured. On arriving home, Dr. Schmidt received a report from the Norwegian survey ship S.S. 'Michael Sars' stating that they had captured a smaller leptocephalus further west in the Atlantic. Schmidt then obtained the co-operation of as many Danish fishing vessels as possible, persuading the skippers to keep a look-out for leptocephali by casting special nets at points further west in mid-ocean. Although 'only one hundred and twenty tiny leptocephali were caught, by plotting their position of capture on a chart, an interesting picture was obtained. This showed that the leptocephali travelled along definite migratory routes — they did not occur haphazardly over the north Atlantic. In 1913 Dr. Schmidt undertook an oceanographic expedition aboard the schooner 'Margarete'. From all the information obtained on this voyage and from previous records he realized that the size of leptocephali decreased the further west in the Atlantic that they were captured. Their travel routes were confined to the waters of the Gulf Stream and the spawning grounds were deep down in the



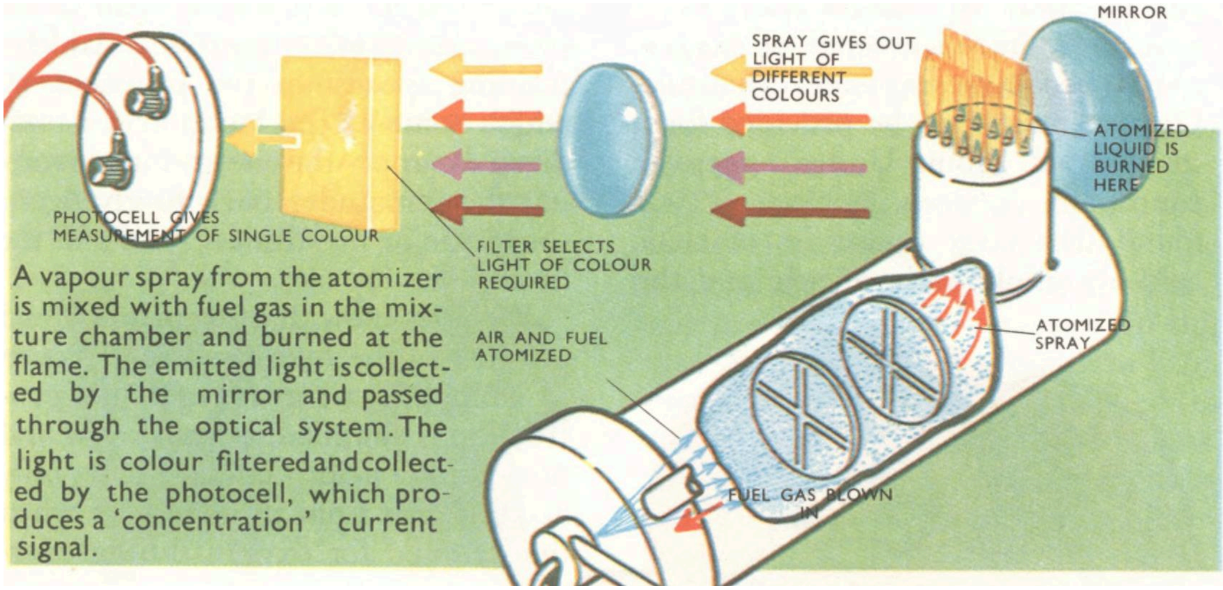
Sargasso Sea, an area rich in floating seaweed south of Bermuda; at least this was the area where the smallest larvae (only $ inch long) were found. Contemporary with Schmidt's work, the studies of naturalists in America had shown that American eels (which closely resemble European eels) also leave the rivers of the Atlantic coast for the Atlantic to spawn. The young are superficially identical and could only be told apart from the number of body segments each possessed. The majority of American cel larvae have between 106 and 109 segments, the European eel between 113 and 117. In the Sargasso spawning ground there appeared to be two hatching areas, one for the American species, the other for the European species. But this aside, the pattern of the European eel's life cycle seemed clear. The adults left the European rivers in the autumn and travelled westward for some 3,000 miles, presumably arriving around the end of the year since early in the following year the first larvae are obtainable. These were captured at around 1,000 feet and this is probably where spawning takes place. These 1,000 feet larvae are around } inch long (7 mm.). Their length is quadrupled in the first summer by which time they have travelled 1,000 miles carried in the current of the Gulf Stream. For two more years they 'ride' the Gulf Stream, reaching the length of around three inches before changing into elvers which make for the rivers of Europe. The American Eel has a much shorter journey, only a third that of its European counterpart. They grow more quickly than the European species, reaching a length of 3 inches in a year. Though the eel's story might appear to be settled once and for all, there is still a certain amount of controversy. Various facts have still to be explained. No adult eels have been caught in the region of Sargasso Sea. In fact, none has been caught further west than the Azores. Specimens caught near the European coast of the Atlantic have had little in the way of fat reserves. It is unlikely therefore that an adult European eel is able to swim against the current of the Gulf Stream 3,000 miles to the Sargasso Sea unless they swim at greater depths where the effects of slower moving deep currents are lower. After all, an adult salmon, in similar conditions, travels a journey only of a few hundred miles, certainly not thousands and many fail to reach the gravel beds where they are to spawn. If the European eels do travel all the way to the Sargasso Sea, then their incredible navigational sense can only be wondered at. No anatomical structure is known. Possibly, as may be the case with salmon, the sense of taste plays a part in ‘sampling’ the quality of the water. If, as is probable, the larvae are distributed at random by the Gulf Stream washing the coast of Western Europe, it is most unlikely that the larvae of Spanish eel parents are going to return to Spanish rivers or of Scottish eels to Scottish rivers. Even more unlikely is it that young eels randomly distributed are going to find their way back in different directions to the same spawning ground where they were hatched. It is just possible that eels from only one part of Europe ever find their way back to the Sargasso Sea. At least one biologist has proposed that stocks of European Eel are maintained from the millions of larvae hatched from eggs of the American species. Those that get caught up in the Gulf Stream are carried eastwards across the Atlantic and distributed by it to the coastline of Western Europe: those that swim northwards reach the Atlantic shores of America. This certainly is a possibility, for it has long been in doubt whether the European and American forms are in fact separate species. Female eels seem to take at least eight years to reach sexual maturity; male eels at least five years (when they are from 3 to 5 feet long). A few individuals never mature sexually and have a longer life — up to twenty years or more and also reach a larger size. Those that do reach sexual maturity are thought to spawn once and then die. “ine eels are bony fish belonging to the family Apodes. Two marine relatives are shown here. (left) A moray eel eating an octopus and (right) a conger eel. The moray is a giant amongst eels often reaching a length of ten feet. ee, ==

still a certain amount of controversy. Various facts have still to be explained. No adult eels have been caught in the region of Sargasso Sea. In fact, none has been caught further west than the Azores. Specimens caught near the European coast of the Atlantic have had little in the way of fat reserves. It is unlikely therefore that an adult European eel is able to swim against the current of the Gulf Stream 3,000 miles to the Sargasso Sea unless they swim at greater depths where the effects of slower moving deep currents are lower. After all, an adult salmon, in similar conditions, travels a journey

BIOLOGY SOME plants are carefully avoided by Man and animals. The reason — they are equipped with tough, sharppointed structures capable of inflicting irritating, even painful, wounds — the prickles and needles and thorns and spines of everyday language. A careful, if cautious, inspection will show that, though often very similar in design, spines in fact have developed from completely different structures. In the Hawthorn tree and the Sloe the sharp, woody spines so unpleasant to gatherers of may-blossom or collectors of sloes, are very small secondary stems — dwarf branches. They develop from axillary buds on an ordinary stem the same as any other branches. But instead of increasing in length and girth, they stop their growth and remain short, pointed structures. Despite their size, however, leaves may be supported, providing urther evidence of their true origins. It is the true leaves in other plants which are modified into spines. The Barberry, for instance, has some of its leaves replaced by long spines. Gooseberry bushes are spinous because of pointed outgrowths from the leaf-base. In the False Acacia the stipules, usually tiny outgrowths at the base of a leaf stalk, are modified into spines. Another possibility is for the margins of the foliage leaves to become drawn out into tapering extensions made rigid by the abundance of mechanical tissue — as in holly and thistle. The decurved thorns of the blackberry and the rose are outgrowths of the stem called emergencies. Unlike true branches, they do not arise from buds in the axils of leaves. Spines covering the cactus are simply outgrowths of the epidermis: they are plant hairs. Gorse — the prickly bush of heathHAWTHORN Hawthorn — spines are really dwarf branches given off from main stems. The spines may support leaves. LEAF. OCOTILLO sae SepNESTAl MID-RIBS GOOSEBERRY SPINES ARE OUTGROWTH OF LEAF BASE LEAVES AND BRANCHES FORM SPINES Modifications of leaves into spines. As a rule, spines are plant adaptations to dry conditions and have only a secondary protective value. THORNS, PRICKLES and SPINES lands, conspicuous by its bright yellow flowers - possesses both leaf and stem spines. The leaf is short, cylindrical and ends in a point. Emerging from the axil of the leaf is a dwarf branch almost exactly identical in appearance, also with a point. To confuse the arrangement even further, the dwarf branch itself may carry more spiny leaves. Functions of Spines Most prickly plants live in dry or THORNS ON STE OF BRAMBLE CACTUS Cactus spines are outgrowths of the plant epidermis. Thorns or emergencies of bramble are climbing devices. moderately dry climates or places where water is soon drained from the soil. Moisture is at a premium — the surface area of the plant over which water can be lost becomes lowered. Commonly this means a reduction in the surface area of leaves, giving them a spiny appearance. Gorse supports this view — if grown in moist conditions few or no spines appear and normal foliage leaves are produced. The disappearance of leaves altogether, with the subsequent development of other photosynthetic structures is also an advantage in dry conditions. Such structures, usually possessing far more mechanical tissue than leaves, are not so likely to shrink on drying. The development of other types of spine commonly seems related to a plant's tendency to increase its woody tissue in times of low water supply but certainly once formed, the woody spines do afford protection. The thorns of bramble and rose, however, have nothing to do with conditions of drought — their primary function is to assist climbing.

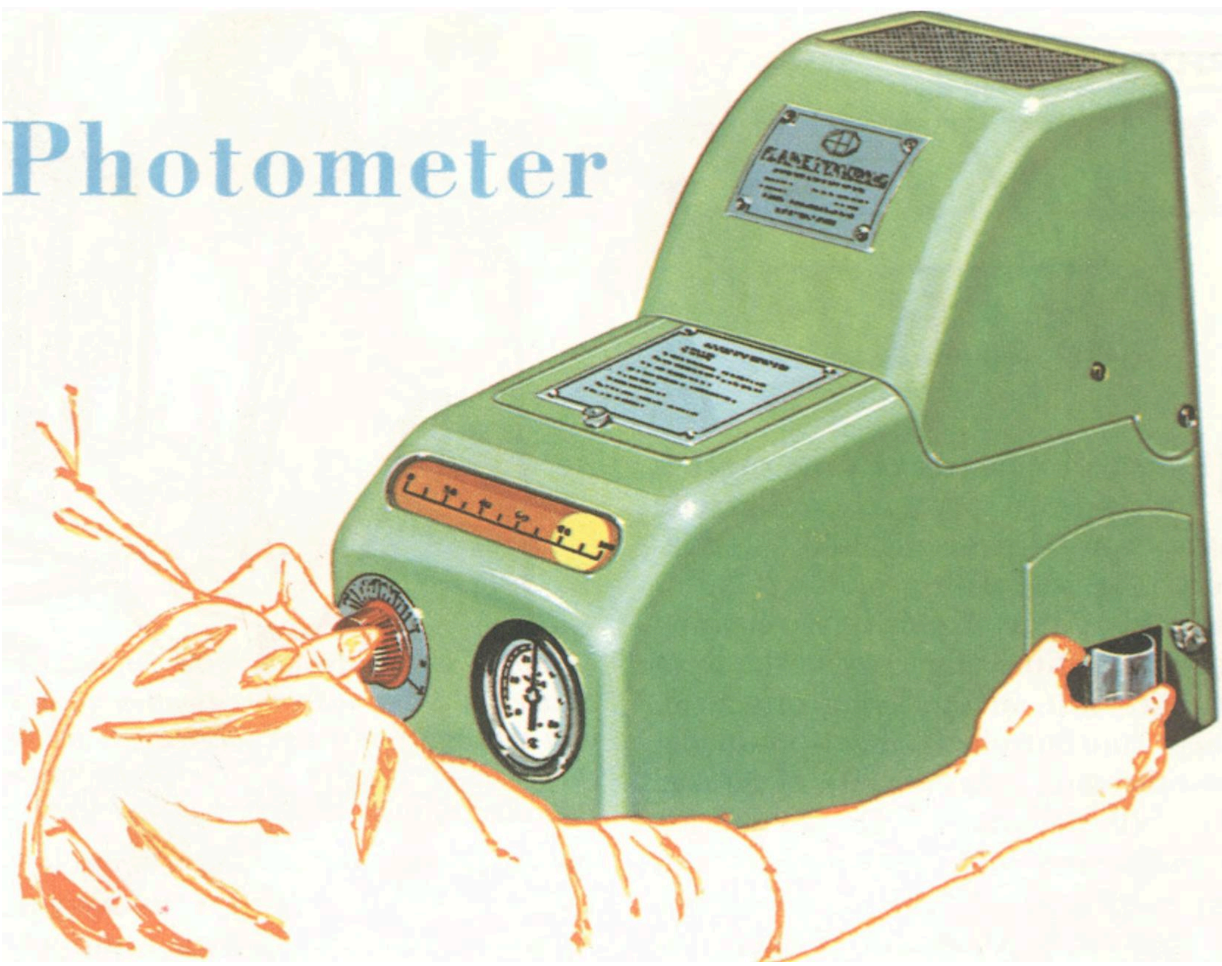

DWARF
BRANCH
HAWTHORN

The Flame N the analysis of inorganic chemicals, one of the simplest but most positive forms of identification is the flame test. The unknown compound is dissolved in acid and a small quantity is picked up on a clean platinum wire. The wire is then held in a flame. If one of a number of particular metallic elements is present, the flame immediately takes on a distinctive colour ~ yellow for sodium, crimson for strontium, green for barium. This simple method is astonishingly sensitive —a sodium concentration of a few parts per million will give an unmistakable yellow flame, and it is used to detect all members of the group of alkali metals. The light is given out by the flame because electrons of the atoms in the compound are excited by the thermal energy of the flame. The electrons are excited into higher energy electron shells, and emit light when they return to their original shells. Each particular type of electron transition give rise to light of a particular wavelength and atoms of different elements give out a particular set of different wavelengths. The result is that each element gives out a unique, easily recognised mixture of light radiations. The simple flame test provides useful results in the laboratory, but the method can be improved upon to give even more valuable information. The simple test tells the skilled analyst, with some degree of certainty, what metal the compound contains, but it does not tell him. how much it contains. The kinds of atoms present determine the colour of the flames, the Using a flame photometer. The solution containing e metal radicals is inserted in the beaker and when the appropriate filter is used the meter reading indicates metal concentration. intensity of the flame is determined by the number of atoms present that are emitting light of a particular colour. So, if it were possible to have a standard, steady flame fed with a steady quantity of solution, the quantity of compound present would be determined by measuring the amount of light of a particular wavelength given out by atoms of the compound. To do this, the light emitted by atoms of the particular compound needs to be isolated and then measured in intensity. This is done in a specially designed instrument — the flame photometer. In the flame photometer the solution is vaporized in an atomizer and the droplets are fed into the fuel gas stream and carried to the burner orifice. The fuel gas can be coal gas or calor gas, SPRAY GIVES OUT LIGHT OF COLOURS AIR BLOWN IN UNDER CONTROLLED PRESSURE although in some instances hydrogenoxygen or acetylene-oxygen mixtures are used. The droplets are heated until the solvent evaporates, and the dissolved compound vaporizes. When the compound is vaporized, the atoms are excited and emit the characteristic light. A spectral ‘line’ or band of a particular wavelength is isolated from the light given out. In one type of flame photometer, this is done by passing the light through an optical filter. Separate filters are provided for each of the metals to be determined — the ‘sodium’ filter, for example is placed in the path of the light received by the photocell when sodium content is being determined. The photocell converts the light into an electric current that is measured using a sensitive galvanometer, calibrated in terms of sodium, lithium, calcium, or potassium content, depending on the filter used. This method provides a very sensitive measuring technique, used to measure, for example, the minute quantities of calcium in blood serum. Apart from biological applications there are many others — checking the effectiveness of water purification plant is one. Concentrations of less than one millionth of a gram per litre can be accurately measured. 2Y24

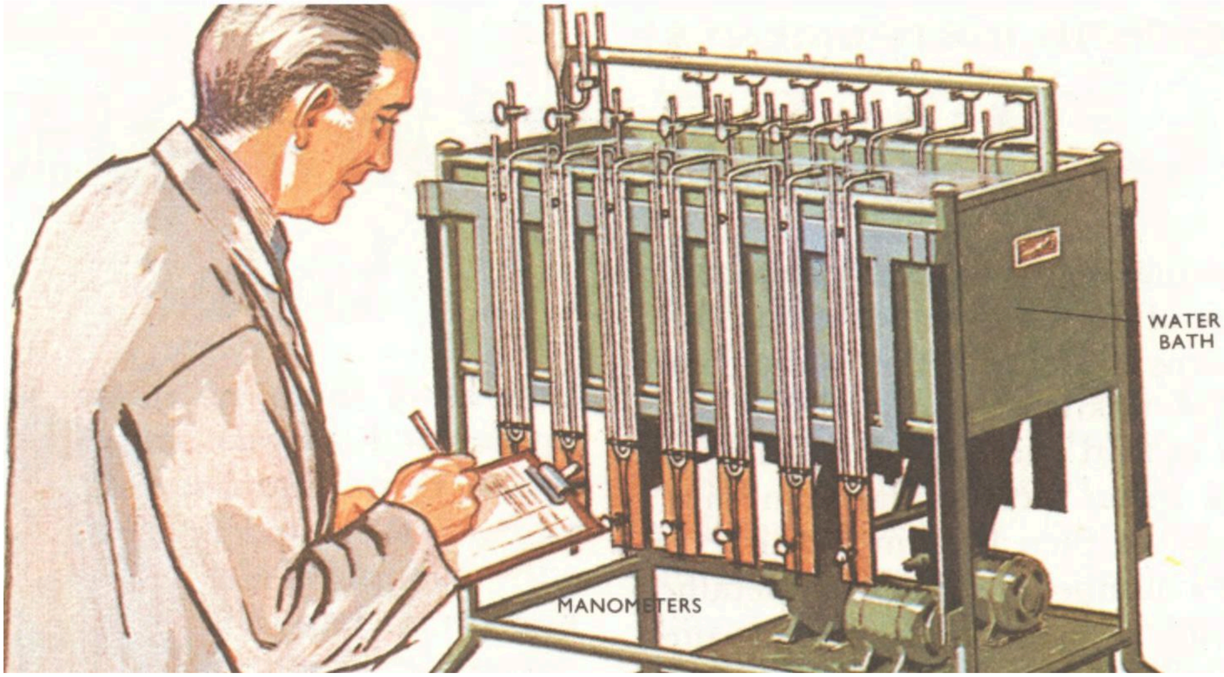


A vapour spray from the atomizer is mixed with fuel gas in the mixture chamber and burned at the flame. The emitted light is collected by the mirror and passed through the optical system. The light is colour filtered and collected by the photocell, which produces a 'concentration' current signal.

Photometer



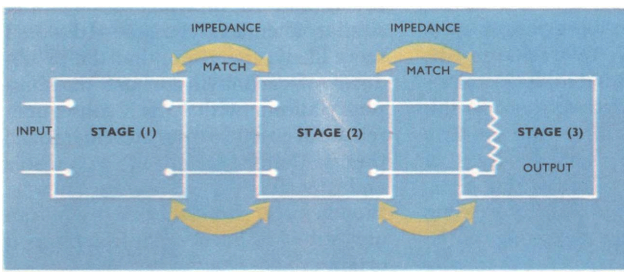
PHYSIOLOGY LIVING 'FIRE' IF a small quantity of cane sugar is placed on a spoon, a little cigarette ash is sprinkled over it and the mixture is ignited with a match it burns fiercely, almost explosively, releasing large quantities of energy in the form of heat. But, in the living organism, such uncontrolled, heat-producing (exothermic) reactions are of no use. The cell machinery is far too delicate to withstand high temperatures and such vigorous reactions would destroy it. The principal source of energy in the body is sugar (glucose rather than cane sugar) and the energy is released from it in a controlled step-by-step manner so that the cells do not blow themselves to bits or shrivel up through producing too much heat too quickly. If the experiment with the sugar is repeated without adding cigarette ash, little or no burning is observed. The ash has a catalytic action in some way which speeds up the chemical reaction considerably. The chemical reactions within the body cells by which sugar is 'burnt' to release the energy needed to drive various other metabolic processes are also controlled by catalysts called enzymes. Some of the chemical pathways by which glucose is 'burnt' or respired are known. This has been elucidated primarily by a painstaking series of experiments with 'juices' extracted from muscle tissue. Though the results of such experiments do not prove conclusively that this is what happens in living muscle -—since chemical and physical distortion of the cell components must have occurred in obtaining extracts—it is very strong evidence. Associated with these extract experiWarburg manometers ake used for measuring the gaseous metabolism of isolated tissues. Saline and tissue under investigation are placed in a special flask immersed in the water bath at a set temperature and connected to the manometer. Readings on the manometers indicate the gaseous exchange of oxygen and carbon dioxide. Analysts of the flask contents also indicates chemical changes that have occurred. ments are others on whole muscle — measurement of changes in oxygen uptake, carbon dioxide output, changes in the concentration of food substances such as glycogen, glucose, fatty acids and so on ~ all help to complete the picture of the metabolic activity of the tissue. Radio-active isotopes are also used as_ tracers. Coupled with these chemical techniques are physical methods of analysis which do not distort the cell chemistry in any way. These include the measurement of heat production, pH (acidity), electrical changes and volume changes. The reserve of carbohydrate in muscle is glycogen or animal starch, whose large molecule is composed of glucose units. Pieces of muscle can be isolated and made to contract by electrical stimulation. | Chemical analysis then shows that the store of glycogen has become depleted, oxygen has been consumed and carbon dioxide liberated. However, a muscle can be made to contract many times repeatedly in the absence of oxygen, until it becomes completely exhausted. Oxygen must then be provided for it to contract again. Under anaerobic conditions (oxygen lacking) it was found that large quantities of lactic acid accumulated in muscle and this also happens even during severe aerobic conditions. The basic source of lactic acid is glycogen and it is produced by the hydrolysis of glycogen through enzyme action with the production of energy as heat and for doing mechanical work. The lactic acid disappears from the muscle when oxygen is admitted and heat is produced. The energy released by the oxidation of one fourth of the accumulated lactic acid is sufficient to reconvert the remaining four-fifths back into glycogen. In this way there is a readily available store of energy for immediate use, and after exercise the store can be recharged by slower, less violent oxidation of lactic acid. This briefly was our knowledge of the chemistry of muscle contraction around 1920. From more recent work we know that lactic acid is not actually formed during the actual contraction process but afterwards (it is this build up of lactic acid that produces the characteristic 'stiff feeling some hours after taking violent exercise). Athletic training speeds up the processes by which some of the lactic acid is oxidized to supply the energy for converting the remainder back to glycogen; thus little or no stiffness results in the fit athlete. We now know that the primary end product of the hydrolysis of glycogen by which the energy for muscular contraction is supplied, is pyruvic acid and not lactic acid. Lactic acid only accumulates from pyruvic acid when the demand for oxygen during exer


WATER
BATH
MANOMETERS

cise is greater than the supply — that is under anaerobic conditions. The change from glycogen through glucose to pyruvic acid takes place in eleven distinct steps. Each step (except one which proceeds of its own accord) is controlled by an enzyme. Firstly the links between the 'glucose' units that make up the glycogen molecule are broken by hydrolysis and each unit is phosphorylated — i.e. has a phosphate group added to it-—to glucose-1phosphate. Via a succession of such phosphate compounds eventually pyruvic acid is formed. Under aerobic conditions pyruvic acid is quickly 'absorbed' into another process — the citric acid cycle. This cycle, besides removing a waste product, also supplies hydrogen necessary for the reduction of certain cell constituents which have been oxidized while taking part in the hydrolysis of glycogen and _ other processes. Adenosine triphosphate (A.T.P. for short) is a vital substance in the breakdown of carbohydrate. The breakdown of glycogen to glucose units under the influence of an enzyme takes place only in the presence of A.T.P. This donates phosphate to glucose thus yielding glucose-1-phosphate — itself becoming A.D.P. (adenine diphosphate). In other reactions A.D.P. accepts 'phosphate' becoming A.T.P. again. The energy for muscle contraction ultimately comes from the breakdown of glycogen to pyruvic acid and the donation of phosphate from A.T.P. during the steps of this process net only transfers phosphates but in changing to A.D.P. considerable quantities of energy are made available. Much of this is undoubtedly used to An unfit person will feel stiff sometime after a race because lactic acid accumulates in his muscles. An athlete feels little or no stiffness because training has speeded up the processes by which lactic acid 1s broken down or ie to aie > Re GLYCOGEN (only @phinute part of molecule shown) ENZYME: ORYLASE: PHOSPHATESDONATED BY A.T.P. WHI CHANGES TO A.D.P. GLUCOSE-1-PHO GLUCOSE-6-PHOSPHATE OTHER INTERMEDIATE STAGES recharge the muscle which in general terms means re-forming A.T.P. It is thought that another organic phosphate compound — creatine phosphate —1s actively concerned with this using some of the energy released by the change of A.T.P. to A.D.P. Creatine phosphate splits, yielding creatine and phosphoric acid, which provides the 'phosphate' in A.T.P. A startling discovery made recently is that the enzyme that catalyzes the breakdown of A.T.P. is in fact the actual protein fibre which makes up the bulk of muscle tissue and which contracts — so causing the muscle to shorten. The breakdown of glycogen provides the energy to recharge the muscle. It is an anaerobic process — one not requiring free oxygen. In this way an oxygen debt of around 20 litres can be withstood. The products of this process, pyruvic acid and lactic acid, are oxidized aerobically yielding energy and glycogen so paying off the oxygen debt. The fact that there is a A SIMPLIFIED Nair ese s OF THE BREAKDOW GLYCOGEN IN MUSCLES TO PYRUVIC ACID AND LACTIC ACID PART OF LACTIC ACID OXIDIZED AND ENERGY USED TO CONVERT REST BACK TO GLYCOGEN LACTIC ACID PYRUVIC ACID CITRIC —> ACID CYCLE limit to the oxygen debt that can be withstood explains why it is only possible to sprint flat out for short distances. A portion of the energy released in muscle contraction is available as heat and it is primarily this source of heat which maintains the temperature of the blood above that of the surroundings. Similar analysis of other tissues shows that the same chemical processes that occur in muscle also occur in liver, kidney and brain cells. Extracts of yeast, the fungus that is commercially important in the making of bread and in brewing, also exhibit similar chemical changes. It would seem likely, therefore, that the breakdown of carbohydrate and its phosphorylation with the subsequent release of energy constitute a metabolic process that is basic to life as a whole and not just to vertebrate muscle. Bio-chemical similarities of this nature are strong evidence for a common origin of life. 2123

| ELECTRICITY | ~ MATCHING HE amplifiers in a radio set increase the volume of sound which is eventually produced by the loudspeaker. They may amplify voltage (the usual way with valves) or current (in transistors) but they are both in fact amplifying electrical power. Power (in watts) is equal to current (in amps) multiplied by voltage (in volts). It is the power of a signal which determines the amount of sound the current can produce when it is fed into a loudspeaker. The output stage of a radio set is called a power amplifier. It is connected, or coupled, to the loudspeaker. A coil in the amplifier circuit and a coil in the loudspeaker circuit are both wound around a magnetic iron core. Together they form a_ transformer, through which they are linked electromagnetically. Fluctuating currents around one coil induce fluctuating currents around the other coil, and the signal current is transferred. Why is a transformer necessary? The reason is that it enables power to be transferred across to the loudspeaker efficiently. There is no point in wasting the power given to the signal during amplification. This is what would happen if the loudspeaker were just connected straight into the output circuit. Direct Coupling The loudspeaker itself has a very low electrical resistance (between one and fifteen ohms). This is the resistance of the loudspeaker coil (not the transformer coil). The current flows around this coil and attracts a piece of iron, so that it moves to-and-fro with the fluctuating current. The iron is coupled to the loudspeaker cone, which then vibrates to produce sound waves. If the loudspeaker were connected directly into the amplifier circuit, the same electric current would flow around the whole circuit, through the amplifying valve, and then through the loudspeaker coil. The valve itself has a large resistance to outgoing currents — of several thousand ohms. As the current flows across the valve, its voltage drops considerably (the voltage drop implies that work is being done in pushing the current against the resistance). But the voltage drop across the loudspeaker coil would be comparatively small. The voltage drop across any electrical resistance is proportional to the resistance. Large resistance, large voltage drop: small resistance, small voltage drop. By far the largest proportion of the work done by the current would be done in the valve's resistance, and not where it is really wanted —in the loudspeaker. It is most efficient to transfer the power of the signal when the output resistance of the valve is equal to the input resistance of the next stage, the loudspeaker. When Rout = 8,000

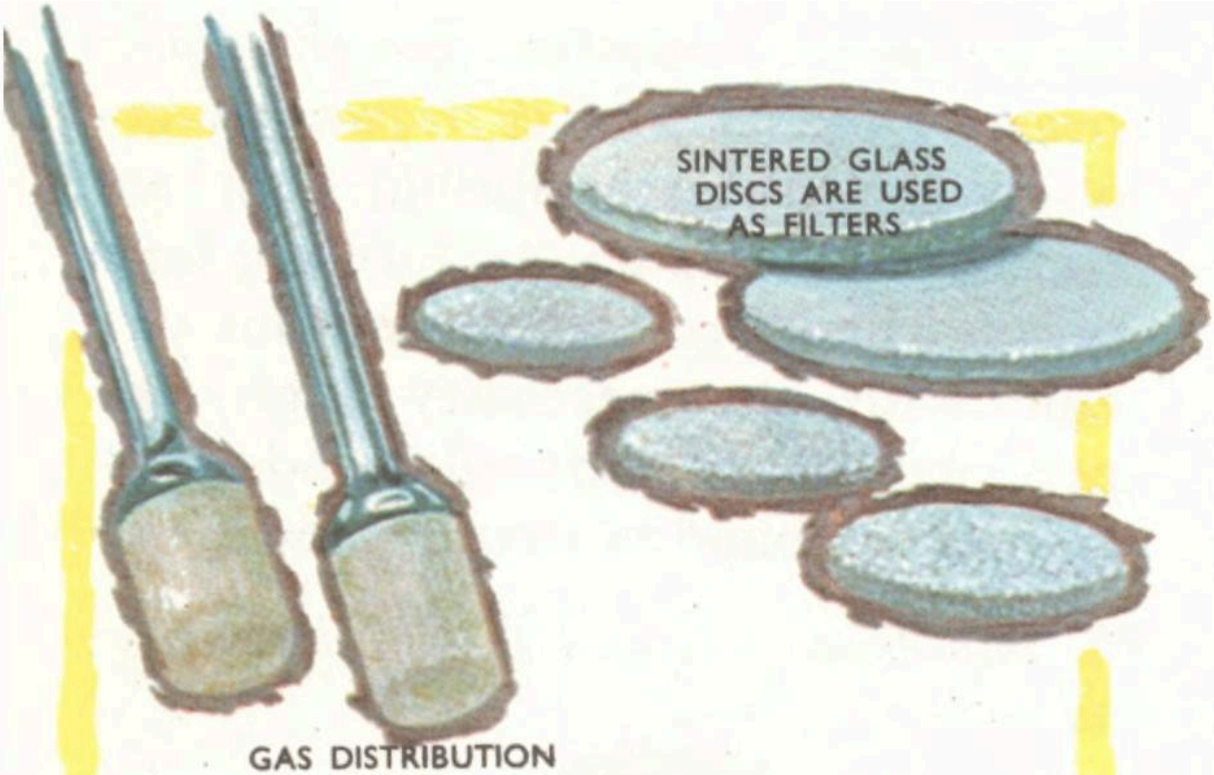
ohms (for a triode) Rour = 15,000 ohms (for a pentode) Rout = around 30,000 ohms (for a transistor) TRANSISTOR The problem in a radio sets to match the output amplifier, with its high impedance . . these two resistances are equal, the circuits are said to be matched. It does not help to increase the resistance of the loudspeaker, because more power is lost in overcoming the additional resistance. Transformer Coupling The transformer puts a different. kind of ‘resistance’ or impedence into the circuit. It is called inductive reactance, - and is a form of resistance to alternating currents. No energy is wasted in overcoming inductive reactance ~ it is temporarily stored in a magnetic field, ready for almost immediate re-use. The coil in the valve’s output circuit is the primary coil of the transformer. The second coil of the transformer is the secondary, and is connected to the loudspeaker coil. Current flows in the primary coil, builds up a magnetic field in the core which links both coils. Current starts to flow in the secondary. But immediately it meets with the resistance-tochange, the inductive reactance of the secondary. An electromotive force (an 2124



Nearly all electronic devices are made in stages, as the electric current undergoes different stages of processing. Whatever the stage, if the maximum amount of power is to be passed along the line, the input impedance of the stage must be comparable with the output impedance of the previous stage. Valves, especially ultra-high-frequency kinds, often need to be matched carefully. A mismatch further down along the line can affect their performance. Inductance and capacitance, as well as pure resistance, all need to be matched.

.. « to the loudspeaker input, with its low impedance. e.m.f.) in the backward direction tends to oppose the current in the secondary. The opposition is reflected back to the primary coil. Associated with the e.m.f. is a back magnetic field, which partially cancels the magnetic field built up originally by the primary current. The primary current appears to be getting nowhere. Certainly it is meeting more impedance than expected. The large voltage difference developed across the primary coil's terminals is a measure of the difficulties the current is experiencing in getting across. The transformer is designed so that this voltage difference (and hence the impedance of the transformer and loudspeaker as a whole) is equal to the voltage difference across the valve (and hence to the valve's impedance), the condition for maximum power transfer. This can be arranged in the following way. The loudspeaker has a known low resistance, of the order of a few ohms. The lower the resistance in the loudspeaker circuit, the higher the current flowing in this circuit (the secondary) —-so the bigger the back e.m.f. and magnetic field which is reflected back to oppose the primary magnetic field. There are fewer turns in the secondary coil. Each turn links several turns of the primary, and any effect in the secondary is automatically magnified, because the voltage developed across a coil depends on the number of turns in it. This is nota match IMPEDANCE SMALL SMALL VOLTAGE DROP LARGER VOLTAGE DROP Here the loudspeaker is inserted directly into the valve circuit. Very little sound will be produced. Valve and loudspeaker are mismatched. The turns ratio of a transformer is the ratio of the number of turns in the primary to the number of turns in the secondary. If there are 100 turns in the primary and ro in the secondary, the turns ratio is 10. This turns ratio would magnify the loudspeaker's resistance not 10 times but 100 times. The impedance 'seen' by the current flowing in the primary coil is equal to the loudspeaker's resistance (or its impedance), multiplied by the square of the turns ratio. A loudspeaker with a resistance of 10 ohms would reflect an impedance of 1,000 ohms (i.e. 10 x 10% ohms) to the transformer primary. By selecting the right turns ratio, the apparent impedance of the . This isa match IMPEDANCE SAME AS VALVE OUTPUT IMPEDANCE TWO IMPEDANCES MATCH The usual circuit. A few coil turns in the loudspeaker circuit linked with more turns in the valve circuit, through a transformer. loudspeaker can be made comparable with the valve's output impedance. It does not take a very high turns ratio to turn a few ohms impedance into a few thousand ohms impedance; a badly matched circuit into a matched circult. Even a simple transistor set will include several transformers to couple circuits, and match them for maximum power transfer. In the radio frequency and intermediate frequency stages of the set, the transformer coils are connected to capacitors, forming tuned circuits. This presents the output circuit with the necessary high impedance at one particular radio frequency, or band of frequencies. Impedance matching through a transformer. The impedance presented to the alternating current in the primary is magnified by the turns ratio of the transformer. OUTPUT AN ACTUAL TRANSFORMER CURRENT FLOW dé IN THIS TRANSFORMER, PRIMARY OF TRANSFORMER IN VOLTAGE, BUT A ‘STEP-UP’ | CURRENT LOUDSPEAKER SECONDARY OF TRANSFORMER MAGNETIC FIELD CAUSED BY CURRENT BUILD-UP IN SECONDARY "ae FLOW

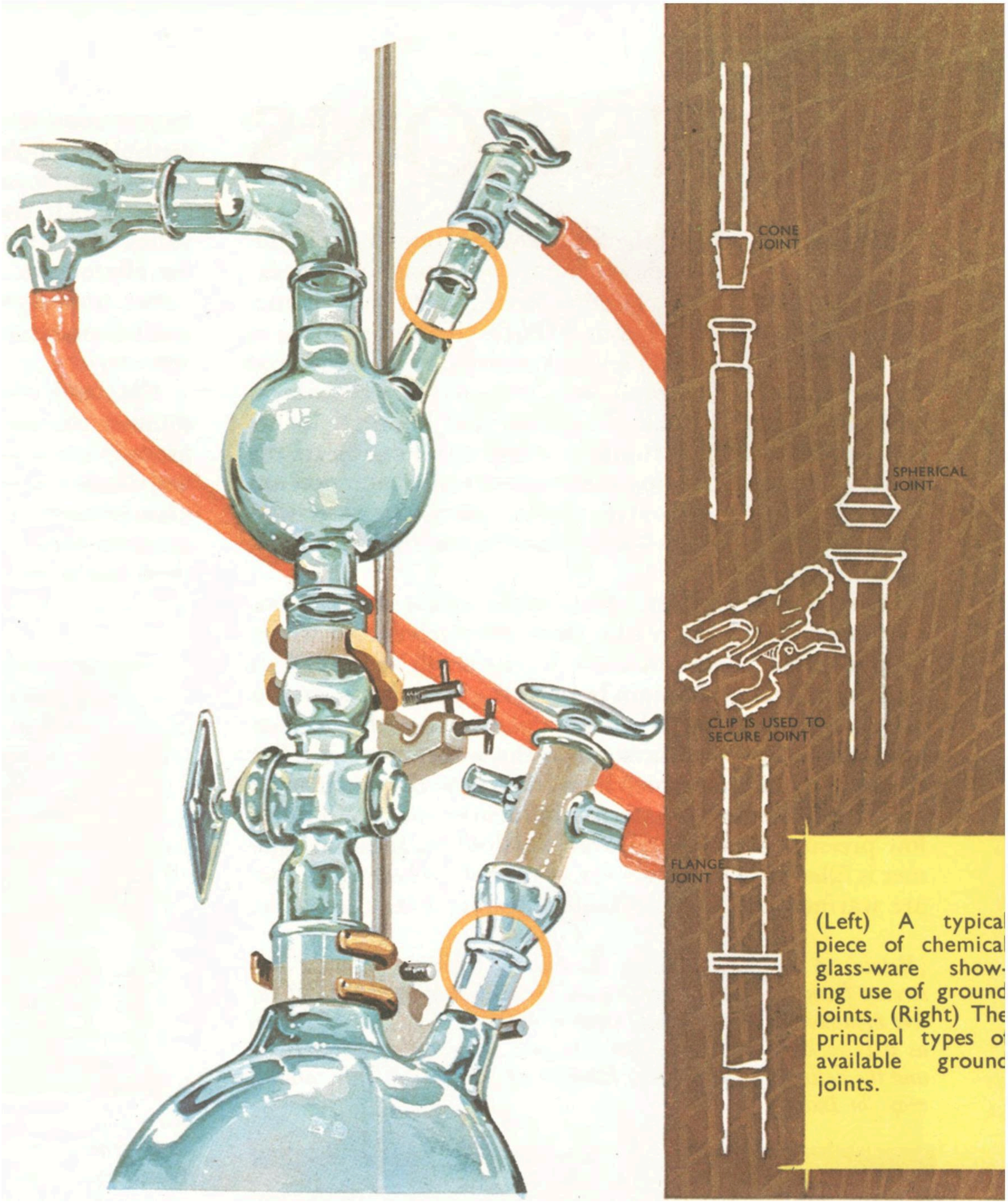
TECHNOLOGY Laboratory Glassware is an interesting fact that, in spite of all the discoveries of modern science — the development of remarkable new materials in particular — the research scientist is still very dependant on apparatus made from one of the most traditional of all materials, glass. Furthermore, there seems to be very little possibility that glass will be replaced in the foreseeable future. Although the basic materials and techniques have remained unchanged over the years, there has been constant enlargement in the varieties of glass that have become available, as well as in the types of factory-produced components. Both in the apparatus factory and in the workshop, the manufacture and construction of glass apparatus have evolved as a pleasing mixture of art and technology. The Nature of Glass A glass is a material which, on cooling from the molten state, does not crystallize out. The result is an amorphous solid that retains some of the properties, transparency in particular, Sintered Joints Another glass component that has come into wide use in the laboratory is the sintered glass filter. This is made of powdered glass fused together to produce a porous solid. They are manufactured in the form of a flat disc or sealed on to the base of thimble funnels. There are several grades of porosity — the finest will filter out particles 2 microns ‘saga cm.) or less in diameter. Sintered glass is used in the bulbs of © gas distribution tubes. These are used to produce streams of gas bubbles in liquids and the bubble size is deterstg by the porosity of the delivery ulb. of a highly viscous liquid. Silicon dioxide (silica SiOz) possesses these essential glass-like properties, but its melting point is too high for it to be used except (as fused silica) for ve special applications. To lower the melting point, sodium nitrate or sodium carbonate is added, but this mixture does not produce a satisfactory glass — it has a fairly high solubility in water, for example. This problem can be overcome by adding calcium carbonate, and the resulting soda glass is a good, general purpose material used in the manufacture of everyday glass objects — jam jars etc. However, laboratory glassware needs to have special properties — having very high chemical stability and resistance to thermal shock. This is provided by the borosilicate glasses (marketed as ‘Pyrex’) and these are made from a molten mixture of aluminium oxide (Al,O3), boron oxide (B,O3), boron, soda and silica. These glasses are probably the mostused and most satisfactory generalpurpose laboratory glassware materials. They are especially valuable because they are free of even the minutest trace of metallic impurity. The presence of trace elements in glass apparatus would have profound effects on some chemical reactions. Manufacture of Laboratory Glassware In the factory the raw materials are melted down in large heated tanks. The melt is then drawn off and used in the manufacture of the various pieces of apparatus. In the tube-making process the liquid glass flows out of the tank to a variety of tube-making machines. Large-bore tube is made in down-draw machines — the liquid melt flows down around a hollow metal cylinder where the tube is formed. Smaller bore tubing is formed by drawing out the melt between rollers and running it away ina continuous process. The small-bore tube may then be re-formed to provide high precision bore tubing. Such tubes are used in the manufacture of pipettes and burettes. Making a chemical condenser. The manufacture of scientific glass apparatus depends largely on the skill of the glass-blower. Beakers and flasks are made using automatic machines. Gobs of molten glass are fed into hollow blanks where the components are formed by plungers or compressed air. Many glass-ware products are made by handblowing the glass. The molten gob is handled on the end of a blow pipe and is blown roughly to shape. The process is completed when the glass piece is placed in a mould where it is blown out to its final shape, fitting the shape of the mould. The most skill-demanding manufacturing process of all is probably lamp working. The glass apparatus is made to a high degree of precision using a high temperature flame and a blowpipe. The: accuracy of the final result depends entirely on the glass blower's skill — all measurements are made with simple instruments like calipers. Annealing Glassware As molten glass cools and solidifies, Master graduation marks on measuring cylinders are made at the surface level reached by an accurately known volume of distilled water. & ™ t Mead aes |



Sintered Joints

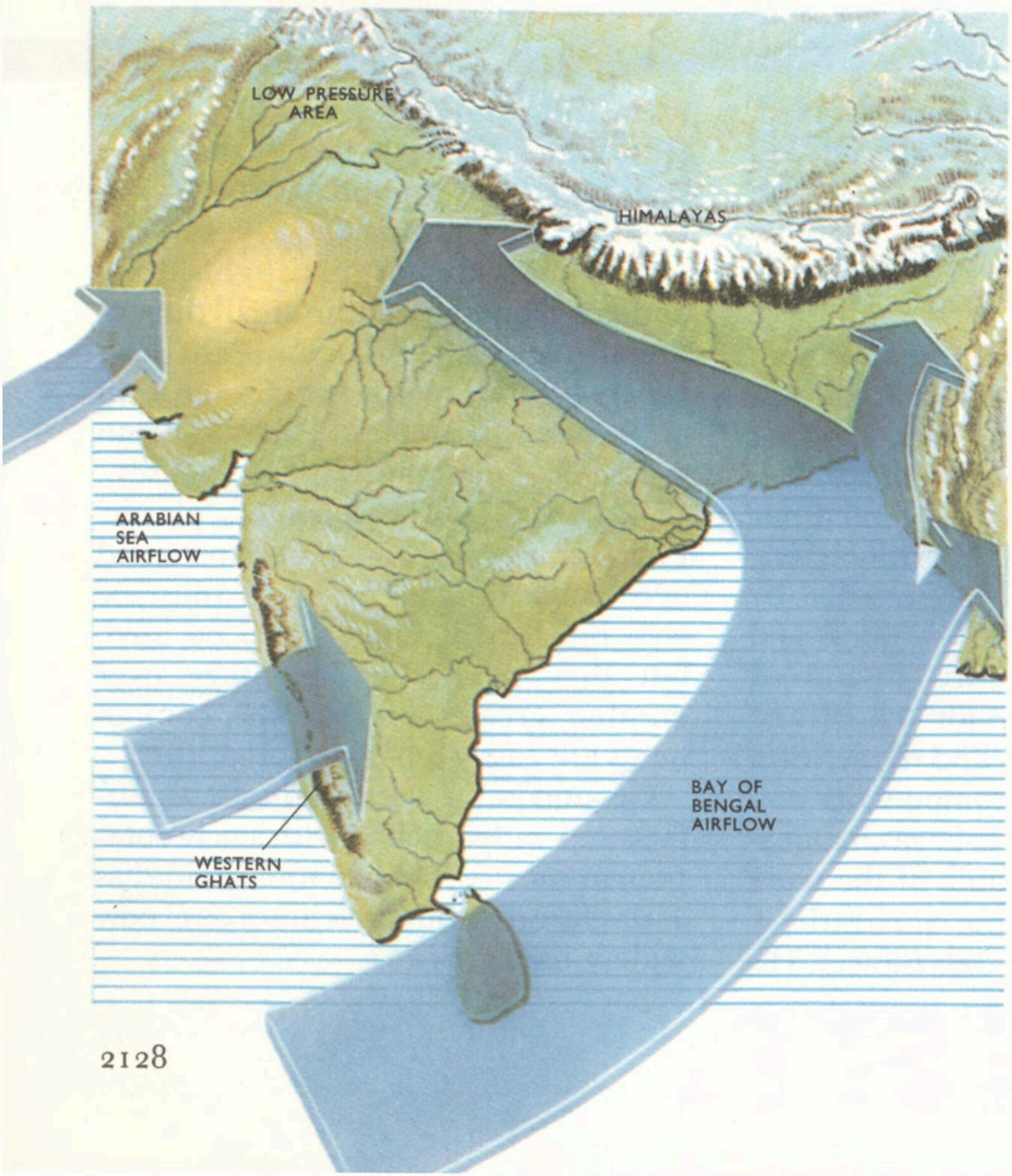
Another glass component that has come into wide use in the laboratory is the *sintered glass filter*. This is made of powdered glass fused together to produce a porous solid. They are manufactured in the form of a flat disc or sealed on to the base of thimble funnels. There are several grades of porosity – the finest will filter out particles 2 microns ($\frac{1}{5000}$ cm.) or less in diameter. Sintered glass is used in the bulbs of gas distribution tubes. These are used to produce streams of gas bubbles in liquids and the bubble size is determined by the porosity of the delivery bulb.

mechanical stresses are set up. If they were not removed they would represent weak points in the glass apparatus. Stressing is relieved by heating the glass object to temperatures just below melting point and then allowing it to cool slowly to room temperature. This process of annealing is carried out in long ovens (lehrs) in which there is a gradual change in temperature from one end to the other. The glass pieces are placed on a moving belt which travels slowly along the oven from the hot end to the cool end, taking perhaps two hours for the whole journey. Graduating Glassware The analytical chemist, as well as other kinds of scientific worker, is absolutely dependant on accurately graduated scientific glassware. Quantities of liquid must, as a matter of course, be measured to the nearest o-1 ml, and conventional equipment (burettes, pipettes) is designed to do this. At the factory, the vessel is first graduated for the maximum graduated capacity by filling it with exactly that volume of distilled water. A mark is made at the level reached. The component is dipped in molten wax, so that an even coating of wax covers the surface. The various graduations are often made on the surface using a dividing tool with the 'master' mark as the reference. A stylus cuts the marks into the wax. The glass component is immersed in a mixture of wetting agent, hydrofluoric acid and sulphuric acid. The graduation marks left in the wax are etched in the glass surface, The graduation marks are scratched in a thin wax layer on the glass surface. The marks are then etched in, using a hydrofluortc acid etching solution. STYLUS MAKES LINE IN W. ON ROTATING G UBI and the vessel is washed and dewaxed. The surface is covered with enamel which is wiped off leaving enamel in the etched scratches, to be fixed by firing in a kiln. Ground Joints One of the more inconvenient features of scientific glass apparatus has always been the various stoppering and jointing arrangements employed. Rubber and cork bungs, bored for inserting tubes and thermometers, have now been largely superseded by ground glass ('Quickfit') stoppers and joints. Complete, easily assembled apparatus is available for school and industrial laboratories and the series of standard joints and components employed permit complete interchangeability. However, the bulk of the work of constructing scientific apparatus is still carried out by the glass blower. He uses the individual conical joints (cones and sockets), ball joints (balls and cups) and ground flanges to put together the equipment. The joints are normally tight-fitting and are ground to within small dimensional tolerances. Vacuum-tight sealing is obtained by the use of some form of grease (petroleum jelly, vacuum grease or sometimes, silicone greases). Solid seals, in the form of 'plastic' P.T.F.E. sleeves, are sometimes fitted over cone joints. The joints must be well matched if these are to be used, but they have the advantage over jointing greases in that the components are easily separated after use. Ground joints were originally manufactured individually with each component formed, shaped and ground as a single unit, but machines are now widely used. 2127

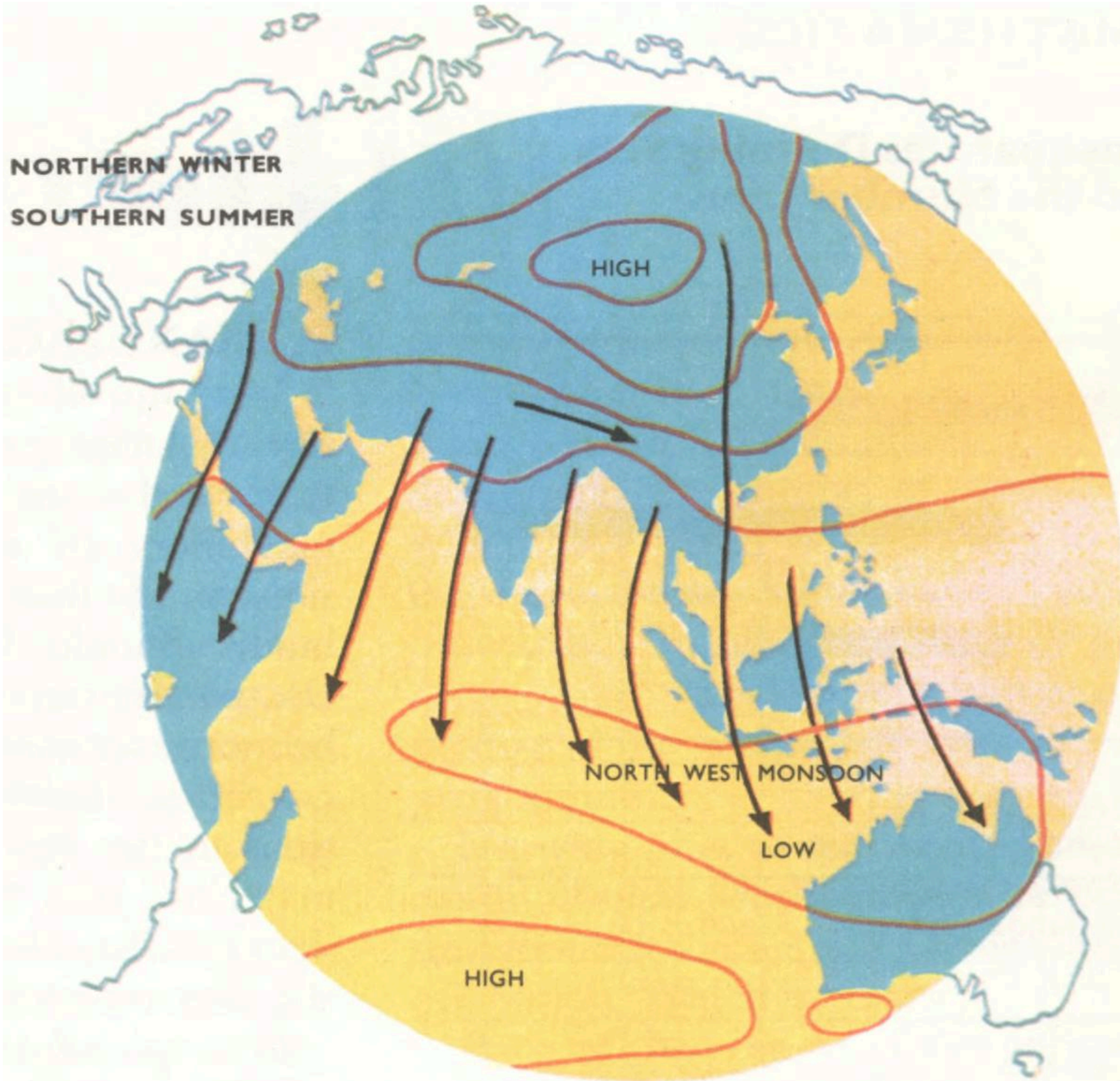


(Left) A typical piece of chemical glass-ware showing use of ground joints. (Right) The principal types of available ground joints.

METEOROLOGY MONSOONS “HE mere mention of the word monsoon conjures up in the mind torrential downpours, accompanying floods, swirling muddy water, families fleeing from their homes for the safety of high ground. ‘What is the wettest place in the world ?’ is a common general knowledge question, and the fact that it is Cherrapunji at the foot of the Himalayas with an annual rainfall of around five hundred inches explains why one also thinks of India when one hears the term. Although the tropical downpours of a monsoon are certainly very spectacular not all monsoons are wet. Monsoon in fact means season; there are both dry and wet monsoons. India and South-East Asia form a region where monsoon conditions most readily show themselves and where ideal monsoon conditions exist. To the north lie the earth’s most formidable mountain barrier, the Himalayas, and its associated ranges, and beyond this stretches the greatest land-mass — Asia; to the south is the warmest ocean. The chains of mountains which hem India in to the north are so effective a barrier to external influences that the intense low pressure area formed over North-West India in summer is filled by winds from the south only. Thus winds that are warmed and enriched with moisture from the Indian (Below) A diagram showing the course of the warm-moist airstream from the Indian Ocean over India during the height of the wet monsoon. Note the great sweep of the Bay of Bengal airstream as it is sucked into the low pressure area over north-western India and the deflecting effect of the Himalayas. (Right) A July rainfall map for India. ‘= on ea en 2 { Y. Ocean roar in over the parched country producing incredible rainfalls over much of the land, particularly along the western coast (due to the influence of the Western Ghats which rise up to 8,000 feet) and up the Ganges valley where the winds are channelled by the solid wall of the Himalayas. But monsoons are not just restricted to India. What conditions are necessary for their occurrence and how do they occur? The land absorbs heat faster than the sea. Hence in summer the continuous heating of the earth’s land masses produce convection currents as the warmed air rises. The net result is a relatively low pressure area over each land mass into which air tends to flow from the relatively high pressure area over the sea, so that winds are created. If these winds are moisture-laden then rain will fall over the land onto which they blow-—this is a wet monsoon. Conversely, the land gives up its heat faster than the sea, so in winter when the air over land masses is becoming cooler, it tends to sink and creates zones of relatively high pressure from which winds blow outwards, that is from the land to the sea. Having passed over land only, the air masses are relatively dry and little or no rain falls — these are the characteristics of a dry monsoon. Any land masses that are large enough to influence the pressure systems in this way will have monsoonal tendencies. A monsoon may be wet or dry depending on the direction of the wind system. An important characteristic of monsoon regions is that the wind direction is reversed from summer to winter and hence there is a marked wet and dry season. A monsoon is in fact an example on a tremendous scale of the natural tendency for winds to blow onshore in the daytime as cooler air blows in from the sea to replace that rising from the land. At night the position is reversed, for the sea retains its heat longer than the land and cooler air from the land blows offshore to replace the warmer air rising off the sea. Areas with monsoonal tendencies are more dominant in the northern hemisphere than in the southern one because a much greater proportion of the earth's land surface occurs in the former. Thus monsoons also affect East Africa

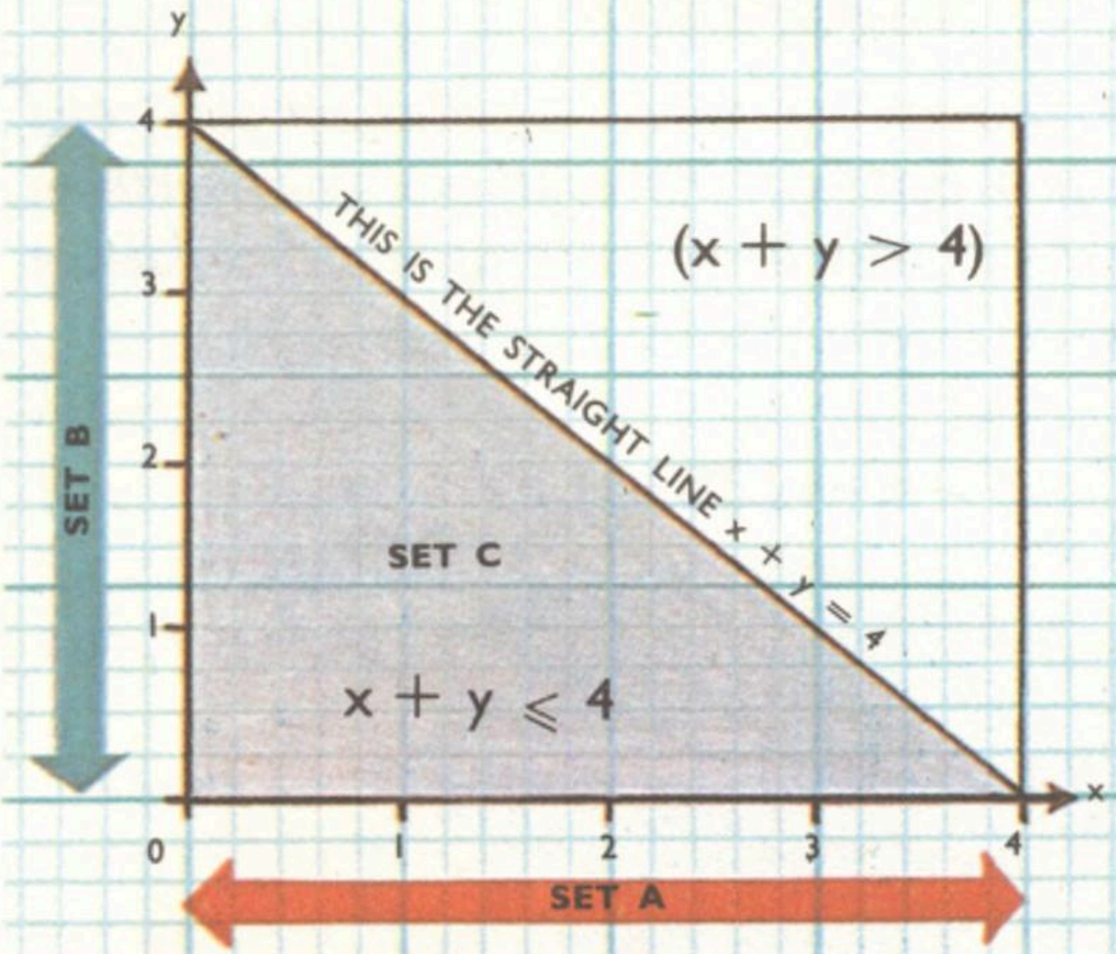

LOW PRESSURE AREA
HIMALAYAS
ARABIAN SEA AIRFLOW
WESTERN GHATS
BAY OF BENGAL AIRFLOW

and Abyssinia, Australia, Europe and North America also show monsoonal tendencies, though the two latter regions have no monsoons as clearly defined as those of other areas. The monsoons affecting Abyssinia, East Africa, Australia, India, South-East Asia, China and Japan are all offshoots of the same great air circulation over and around the Indian Ocean. Pressure over Asia is low in summer when it is high over Australia and the southern Indian Ocean. Conversely pressure over Asia is high in winter but low over, and to the north of, Australia. From May to October, winds blow northwards from Australia towards Asia, first as the south-east trades, before they swing to a north-easterly direction, and as the south-western monsoon deflected by the Earth's spin as they cross from the southern to the northern hemisphere. In their passage across the Indian Ocean they pick up large quantities of moisture which they deposit over India, the Philippines and Japan. Abyssinia and East Africa also receive most of their rain at this time. Australia, where the pressure is high, and consequently which has winds blowing outwards from it, has little or no rain between May and October except in the south which comes under the influence of the wet westerlies. But as the sun returns to the southern hemisphere a low pressure area builds up over northern Australia and the East Indies and a high pressure area over Asia. Winds sweep outwards from this over the East Indies towards Australia and gradually the rains spread southwards across northern and eastern Australia. Central and south-western Australia have very little rain at this time. Little of the winds emanating from the Asiatic high pressure zone affect India because of the barrier of the Himalayas, so that although this is called the cold season, India never experiences the bitter cold that is characteristic of Asia to the north, especially China, in winter. A secondary high pressure area is established over the Punjab and winds blowing outwards and southwards from this join with the Asiatic airflow from China. Only Ceylon and the southeastern tip of India have appreciable amounts of rain at this time, from winds which have picked up moisture while blowing down the Bay of Bengal. By the beginning of spring in the northern hemisphere, India's cold spell has become the hot season, and this increases in intensity as the sun moves overhead further north. Due to a rapid increase in convection currents, the high pressure zone breaks down and a low pressure area is established. This build-up of the low pressure area is a time when violent thunderstorms occur and although these may at times do serious damage to crops they also bring much-needed rain to the spring rice crop in Bengal and encourage the first growth of the tea bushes in Assam. This intermediate, stormy period heralds the approach of the wet monsoon which ‘bursts’ with great suddenness over western India as the airflow from the southern Indian Ocean becomes continuous. The ‘burst’ may be followed by brief fine spells before the rains really start in earnest. The incredible thing about the bursting of the monsoon is that it always occurs within a week or so of the same date each year although the actual time varies from place to Diagrams showing the reversal of wind circulation over al around the Indian Ocean — between the southern and northern hemispheres — as the pressure areas there build up and breakdown with the changing seasons. eo SS NORTHERN SUMMER aie) 2 SOUTHER ER ec? place, according to its latitude. A late bursting of the monsoon or an early retreat can spell economic ruin to farmers, many of whom rely upon the rains to fill the storage tanks that supply the crops with water during the succeeding dry season. The south-westerly airflow reaching India forks roughly into two branches, one, the Arabian Sea branch, depositing most of its rain over the Western Ghats, and a second, the Bay of Bengal branch, sweeps over the sea between Ceylon and Sumatra before being drawn northwards and westwards into the low pressure area over northern India, deflected by the Himalayas. 2129


NORTHERN WINTER
SOUTHERN SUMMER
HIGH
NORTH WEST MONSOON
LOW
HIGH

THERN SUMMER
THERN WINTER
LOW
SOUTH WEST MONSOON
SOUTH EAST TRADES
HIGH
HIGH
WESTERLIES

MATHEMATICS Inequalities Defining Sets SetAisO 30 (Tg and linear. The manufacturer might be intending to use two basic materials X and Y, and his problem is to find the best balance between the two. The amounts of X and Y he must use are not fixed definitely, but each must not be more than a maximum value, nor less than a minimum value. In other words, there is a range, or a set of possible values for Y, and a similar set of possible values for X. Values of X and values of Y are plotted along the x- and y-axes of a graph. Now all possible combinations of X and Y cover a rectangular area on the graph paper. This is the first step in linear programming. It has excluded every single combination of X and Y outside the rectangle. In subsequent steps the area of possible solutions is made smaller and smaller by imposing more restrictions on values of X and Y. Although material X and material Y are both used, there are upper and lower limits to the total amount of material, X + Y. If the total amount of material is to be not less than 1 Ib., not more than 3 lb., then Se es me A product with a total of 2 lb. of a Programming a Programme A very simple example of a linear programme. The radio producer has to fill up at least 30 minutes of transmission time, using 4 live performers, and also recorded music. He can have not more than 30 minutes each of ‘live’ or recorded. The cost ‘ceiling’ is £250 — the recordings cost only £5 per minute, while the ‘live’ part of the programme costs £10 per minute. The two variables are ‘live’ time and ‘recorded’ time, and they are mapped to x and y. As an additional condition, the producer must e (i.e. x cannot be less than 4 x 2 = give each ‘live’ performer at least 2 minutes In



Set A is $0 \leq x \leq 4$
Set B is $0 \leq y \leq 4$
Set C is $x + y \leq 4$

Symbols

A < B means B is greater than A, or A is less than B than B.

$\mathbf{B} < \mathbf{C}$ means C is greater than B, or B is less than C.

$\mathbf{A} < \mathbf{B} < \mathbf{C}$

means B is greater than A, but less than C. The pointed end of the symbol points to the smaller quantity.

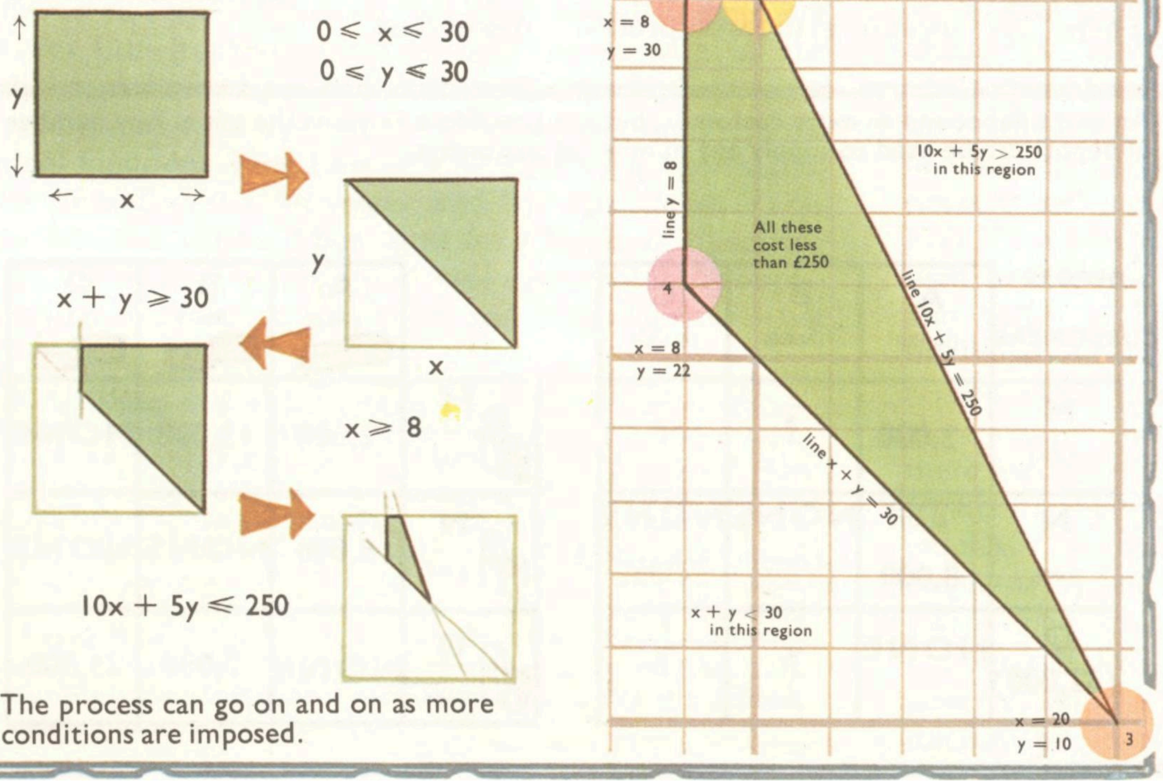
$\mathbf{A} \leqslant \mathbf{B}$ means B is equal to or greater than A.

$\mathbf{A} \leqslant \mathbf{B} \leqslant \mathbf{C}$

means B is equal to or greater than A, equal to or less than C. B is *bounded* by the values A and C.

Linear As more conditions are imposed (symbolized by the shapes) the area common to all shapes diminishes. Narrowing Down the Set of _ Possible Solutions 0< x < 30 0< y < 30 10x + 5y < 250 The process can go on and on as more conditions are imposed. material is acceptable, but a total of 4 lb. is ruled out, and so is a total of 4 lb. Boundaries between possible and impossible solutions are marked by the straight lines: X+ Y=1andX+Y=3. These are parallel to each other, and cut diagonally across the rectangular area. The area bounded by the parallel lines marks the set of possible values for X + Y. It excludes parts of the original rectangle. Next, the relative cost of X and Y is taken into consideration. If the total cost of material is limited to 10/or under, material X costs 5/- per lb., and material Y costs 4/- per lb., then the total cost is: (5X + 4Y) shillings and 5x + 4Y < 10. Again, a straight line, 5X + 4Y = 10 cuts through the area left after the previous step, and excludes even more combinations of X and Y. The process of whittling down the area, and so narrowing the choice, can go on and on, as more factors are taken into consideration. All that is left at the end is an irregularly-shaped area on the graph paper. Its sides are all straight lines, because all the conProgramming ditions imposed on X and Y are linear. The final shape is called a convex polygon. Its number of sides depends on the number of conditions which have sliced wedges away from the originalarea. All the corners point outwards— in other words, they are convex —and this is a very important feature of the shape left after linear programming. Drawing lines to represent limiting conditions has whittled down the possibilities, but there is still a wide choice of solution. Which is best ? Each of the corners of the polygon represents a maximum or a minimum of some form. One corner will represent the dearest product, another the cheapest, or the one with the highest durability or the lowest manufacturing cost. One corner will combine some or all of these to give the best all-round product. The manufacturer has only to look at the corners of the polygon and work out the corresponding cost to see where the best solution lies. Finally, he has to make some choice — though within the narrow limits the linear programming leaves. The Transportation Problem In all linear programmes, the varia

Narrowing Down the Set of Possible Solutions



The process can go on and on as more conditions are imposed.

Programme 1

8 minutes 'live', 30 minutes recorded – i.e. maximum recorded, minimum 'live'. A 38-minute programme for £230, the cheapest cost per minute.

Programme 2

10 minutes 'live', 30 minutes recorded. This is the longest possible programme for £250, but costs slightly more per minute than programme 1.

Programme 3

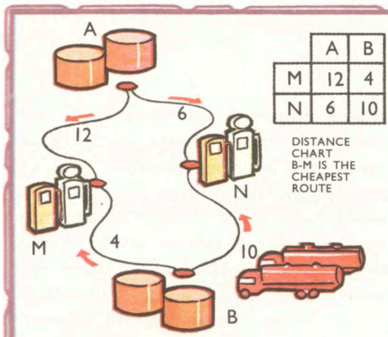
20 minutes 'live', 10 minutes recorded. A 30-minute programme with the highest cost per minute.

Programme 4

8 minutes 'live', 22 minutes recorded

8 minutes 'live', 22 minutes recorded.
The shortest programme for the least amount of money, £190.

DISTANCE CHART B-M IS THE CHEAPEST ROUTE The Transportation Problem — 2 Depots, 2 Customers bles X and Y (and Z in more involved problems) are so simply related that the relationship can be represented by a straight line. Although this is certainly true for another important class of linear programming problem, the transportation problem, the programming is usually done in a table instead of by straight line graphs. Usually there are too many variables to be accommodated on a graph. The transportation problem involves transporting goods from depots to customers. Each customer has certain requirements, each depot has certain stocks, and the problem is to devize the most efficient (usually the cheapest) method of transportation, allotting certain quantities of goods to travel out from each depot to each of the customers. Distances between depots and customers form a separate table, and, in working out the cost of The Transportation Problem — 3 Depots, 3 Customers A-N IS THE CHEAPEST ROUTE DEPOTS A B HAS IN STOCK 10,000 gls. HAS IN STOCK 20,000 gls. CHEAPEST ROUTE CUSTOMERS 15,000 THIS MUST BE 20,000-1 5,000 13 — $8,000 gls. 5,000 On the left is the first step of the programme. The stocks and requirements are marked for each depot and garage and each depot must get rid of all its stock. On the right is the completed programme, which is also the cheapest. any solution, the transportation cost is assumed to be proportional to both the quantity of goods, and to the distance. This assumption is necessary for the programming to be linear. Another table is made to be filled in by goods transported. Depots are marked along the top of the table, customers down the side. Each gap is filled in by the amount to be transported from one depot to one customer. The shortest route is the cheapest, so the first step is to 'guess' a solution, and send as much as possible along the cheapest route. Then goods are allotted to the next cheapest route, and so on. After this the cost of the programme is worked out. It will be a cheap solution, though not necessarily the cheapest, and it corresponds to one of the corners of a convex polygon. There are several standard procex = AMOUNT M GETS FROM A ofa -AMOUNT_M.GETS FROM B IMPOSSIBLE SOLUTIONS IN_THIS REGION — M GETS TOO MUCH M-GETS-TOO LITTLE IN THIS REGION When there are only two variables, it ws possible to solve the transportation problem on a graph. dures for finding out whether the first guess is in fact the best solution. These are all equivalent to going round the corners of a convex polygon until the most economical routing is discovered. An extra depot and an extra customer, but the procedure remains the same. Any number of depots can be used to supply any number of customers. (N NEEDS NO MORE) (N NEEDS NO MORE)



DEPOTS / CUSTOMERS	A (HAS IN STOCK 10,000 gls.)	B (HAS IN STOCK 20,000 gls.)
M (NEEDS 10 – 15,000 gls.)	?	CHEAPEST ROUTE 15,000
N (NEEDS 13 – 18,000 gls.)	?	THIS MUST BE 20,000–15,000 5,000

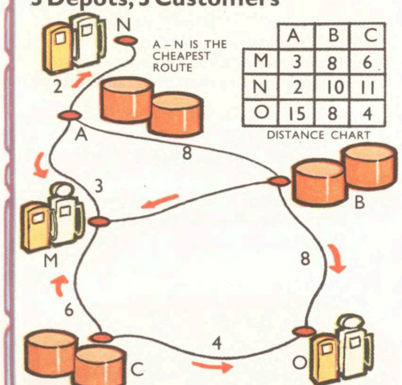
	A EMPTY	B EMPTY
M FULL	NONE	15,000
N FULL	10,000	5,000

The Transportation Problem – 2 Depots, 2 Customers

On the left is the first step of the programme. The stocks and requirements are marked for each depot and garage and each depot must get rid of all its stock. On the right is the completed programme, which is also the cheapest.

The Transportation Problem – 3 Depots, 3 Customers

An extra depot and an extra customer, but the procedure remains the same. Any number of depots can be used to supply any number of customers.



DEPOTS / CUSTOMERS	A (10,000)	B (20,000)	C (25,000)
M (15,000)	2,000	?	?
N (8,000)	CHEAPEST ROUTE 8,000	NONE (N NEEDS NO MORE)	NONE (N NEEDS NO MORE)
O (32,000)	NONE	?	?

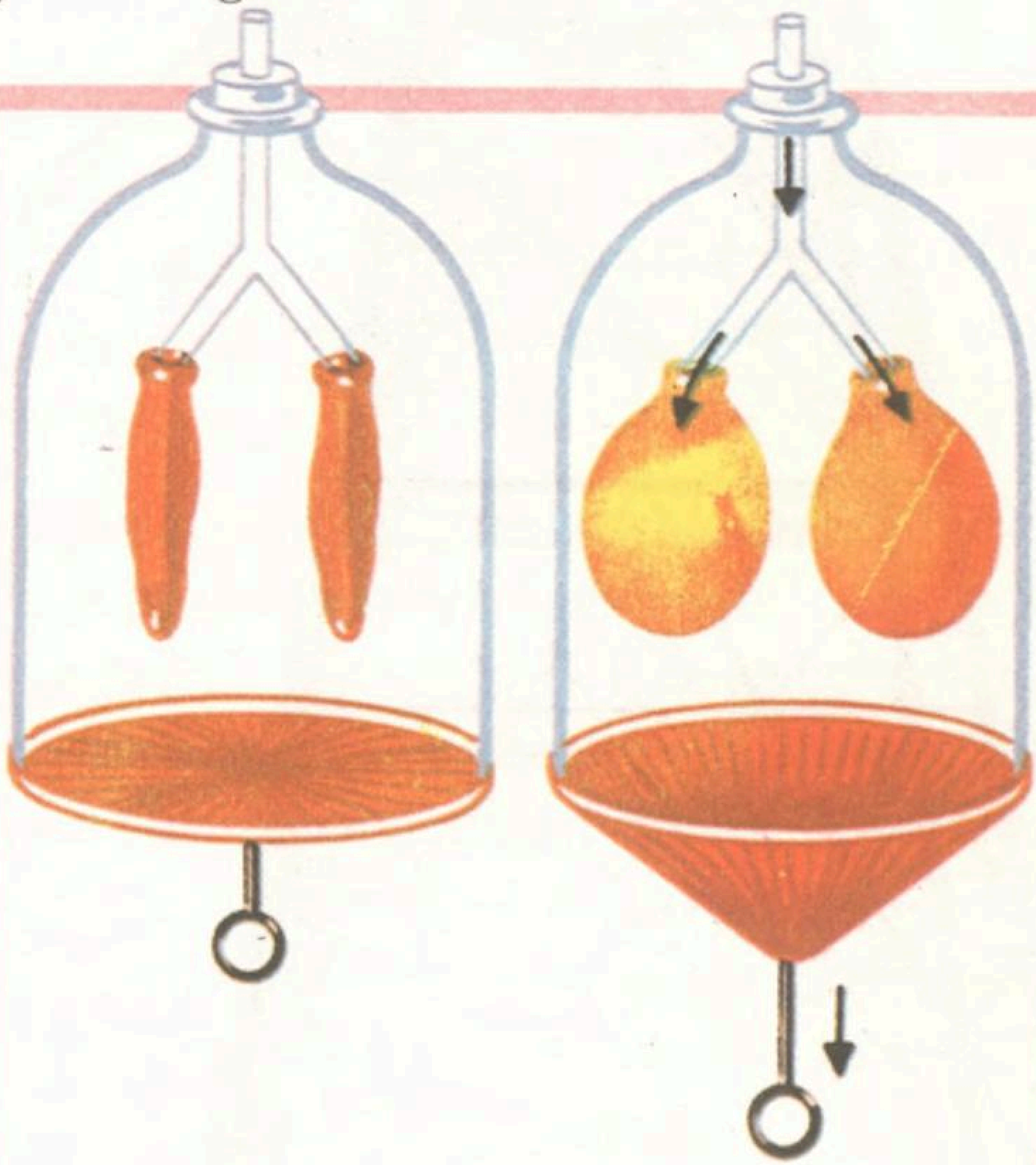
	A EMPTY	B EMPTY	C EMPTY
M FULL	2,000	13,000	NONE
N FULL	8,000	NONE	NONE
O FULL	NONE	7,000	25,000

Lamarek - his Contribution to Science E name of Jean Baptiste Chevalier de Lamarck, French naturalist of the 18th century, is probably better known today than ever before. But usually, he is remembered for a single achievement — as the first man to propound a complete theory of organic evolution. This really does little justice to Lamarck. Most of his long life was devoted to studies in natural history and he emerged as one of the greatest biologists of his age. Lamarck was born in a Picardy village in 1744. He obtained some early education at a college in Amiens but by 17 he was distinguishing himself for valour in the French army. The return to peace in 1762 left Lamarck restless. His temperament was not suited to tedious barrack room duties. He resigned his commission and took up the study of medicine. Supporting himself by working in a bank, he qualified after 4 years. But now, instead of practising his acquired knowledge, Lamarck threw himself into the study of botany. Plants he had always loved and, equipped with a scientific training, he became keeper of the Herbarium of the Royal Gardens. Ten years of hard work ended in the publication of Lamarck's Flore Frangaise — a description of wild plants in France. The book incorporated a key for plant identification which Lamarck had devised himself. A further 15 years of botanical work saw Lamarck, at nearly 50, a leading French botanist. But the really illustrious part of his career was yet to come. In 1793 he was appointed Professor of Zoology at the Paris Museum with special interests in insects, worms, and microscopic animals. Lamarck's small quota of zoological knowledge did not prevent him from pursuing his new course with all the enthusiasm he had devoted to botany. His findings were to revolutionize the systematics — the classification — of the animal kingdom. First Lamarck divided the kingdom into what he called the Vertebrates and the Invertebrates — now very familiar terms. Then after long hours of dissecting in his laboratory he suggested new invertebrate groups — based on anatomical likenesses and dissimilarities — not Lamarck's theory of evolution was the culmination of a life devoted to describing and classifying plants and animals. merely appearances. The group Vermes recognised by Linnaeus, was demolished, Lamarck showing how completely different animals had been wrongly classified together. New groups were made, Lamarck first naming the Annelida, the Arachnida, the Tunicata and the Crustacea. Incidentally, it was Lamarck himself who coined the word biology. From the study of modern invertebrates, Lamarck became interested in comparing them with the remains of past invertebrates. From the Tertiary muds and sands of the Paris basin, he collected a variety of fossils, chiefly mollusc shells. In his writings he encouraged other workers to similarly compare past with present forms. He drew biology closer together with palaeontology and, in fact, has been called the founder of invertebrate palaeontology. In other branches of geology, Lamarck showed a complete mastery of fundamental principles — as is shown in yet another study, his World Hydrology. A man of Lamarck's versatility could not remain unaware of a possible relationship between different forms of life and methods by which one type of animal could change or evolve into another. Stimulated by earlier attempts of Buffon, he presented his own theory of evolution — the Lamarckian Theory. Seeing how closely structures of different organisms were related to modes of life, he stressed the importance of the surroundings — the environment. 'T'wo laws were made. First, that organs continuously used in response to the environment were strengthened, and those not used disappeared. Second, when these modifications have been acquired, they are passed on by reproduction. This second law — the inheritance of acquired characteristics — has met with violent opposition. Certainly there are few if any proven cases of such inheritance. Yet whether right or wrong, this does not invalidate Lamarck's emphasis on the influence of environment. The same emphasis appeared in 1859 when Darwin proposed his Theory of Evolution by Natural Selection. Lamarck died in 1829. A monumental career ended in poverty, and, perhaps due to overwork with lens and microscope, in blindness. 2133

| MEDICINE | "THE simple mechanics of breathing may be explained as follows. The rib cage and its tissues form an airtight box containing the lungs which are in contact with the surroundings by way of the windpipe and nostrils. The base of the box is formed by the diaphragm, a thin, yet tough, domeshaped sheet of muscle. At inspiration (the act of breathing in), the volume of the thoracic box is enlarged as the Breathing Machines arch of the diaphragm becomes shallower due to contraction of its muscle fibres, and as the ribs are raised and moved outwards by the contraction of the intercostal muscles. The increase in volume reduces the pressure on the lungs and so air is drawn into them A ventilator which can be used to supply anaesthetic gases alone or with air in an operation and also air alone in emergency for treatment of respiratory failure as in poliomyelitis. The apparatus applies positive and negative pressures within the lungs. POSITIVE PRESSURE PUMP APPLIES NEGATIVE PRESSURE TO EMPTY LUNGS 2134 LUNGS INFLATED LUNGS DEFLATED Diagrams illustrating the action of the apparatus. through the windpipe and the nostrils —the lungs are inflated. Rapid exchange of oxygen and carbon dioxide is then effected through the thin walls of the alveoli and the capillaries. Expiration (the act of breathing out) is mainly a passive process and air, now richer in carbon dioxide but poorer in oxygen, is forced out of the lungs by the elastic recoil of connective tissue fibres in the lung tissue. The increased pressure on the lungs due to the reduction in volume of the thoracic box as the diaphragm relaxes also helps to empty the lungs. Amongst healthy people there is great variation in the breathing cycle, in the speed of inspiration, its depth and duration, the quantity of air inhaled with each breath, the vital capacity, efficiency of gaseous exchange, the pressures within the lungs and so on. These factors also vary with the degree of exercise — breathing is least active when a person is asleep and most active when he is engaged in strenuous activity. In addition to these normal variations are those resulting from disease — the complete or partial loss of the ability to breathe due to paralysis of the breathing muscles as in poliomyelitis, for example. Obviously in such circumstances breathing must be assisted or completely performed by use of specially designed apparatus. There are two basic methods of approach. With the conventional respirator, or ‘iron lung’ as it is popularly called, the patient is wholly enclosed, except for the head and neck, and a negative pressure is produced within the

chamber at regular intervals by means of a pump, the pressure returning to atmospheric pressure after each negative phase. The negative pressure draws on the chest wall, raising it and so causing air to enter the lungs. On returning to atmospheric pressure the rib cage returns to its relaxed position and the lungs deflate as air passes out of them through the windpipe. Such apparatus is cumbersome and poses considerable nursing problems — access to the cham

PRESSURE WITHIN MACHIN REDUCED SO AIR IS 'DRAWN INTO THE LUNGS PRESSURE RETURNS TO ATMOSPHERIC VALUE SO LUNGS _— ARE EMPTIED A breathing machine which applies a negative pressure to the outside of the chest, thus raising the ribs and causing air to enter the lungs. The patient ts almost totally enclosed within the apparatus. Diagrams explaining the action of the machine. ber is difficult and provided by restricted openings in its walls. At intervals the whole chamber has to be rotated to allow periodic turning of patients who are unable to cough because chest muscles are paralyzed. When complete access to the chamber is necessary for certain nursing procedures the negative pressure mechanism can no longer operate and the airway into the lungs has to be maintained by way of a tube passing through an opening in the neck. However, one desirable feature of this type of respirator (or ventilator) is that its action is to produce actual movements of the chest which simulate those of natural breathing and in doing so exercise paralyzed respiratory muscles. The other alternative to applying a negative pressure on the outside of the body is to apply a positive pressure to actually force air into the lungs. Apparatus which also helps emptying of the lungs by applying a negative pressure after each positive phase is also employed. Such devices are becoming increasingly important in surgery where a breathing machine can wholly take over the body's natural respiratory mechanism — supplying anaesthetic at the same time — so that there are no respiratory movements to interfere with the surgeon's work. This is particularly important in surgery on the chest. They are also important in poliomyelitis when breathing is restricted or impossible on the part of the patient, air being forced down the windpipe into the lungs by way of a tube passing through a hole in the front of the patient's windpipe. Such an arrangement has some advantages over an apparatus that wholly encloses a patient. Nursing staff obviously have easier access to the patient so that washing, injections and other operations can be performed more readily. Psychologically the patient must feel 'freer' than he does within the confines of a large chamber. The design of breathing machines is complicated and requires the careful co-operation of both engineers and doctors. They must be engineered so that they function efficiently from a mechanical viewpoint and at the same time they must be able to satisfy the body's physiological requirements. This means that they must be adjustable over the wide range of breathing requirements of different individuals. During normal breathing, the inspiratory phase is much shorter than the expiratory one so that the pump has to do mechanical work for brief periods followed by longer pauses if it is only supplying a positive or negative pressure, or it has to work for alternatively long and short periods if it is supplying both positive and negative pressure phases. It has to be adjustable so that the ratio of the length of the two phases can be altered. Similarly, the volume must be adjustable over a wide range. Each machine must be able to supply air to a child or a mature adult or even a person with one lung, whose needs will obviously differ considerably. Undoubtedly further refinements will be made in the present breathing machines, but it is to be hoped that this comes about through increased co-operation between scientists of different backgrounds; whether they be trained in medicine, physics, or en_ A model demonstrating the mechanism of breathing can be made using a bell jar with a rubber sheet pulled taut across its large open end. The lungs and windpipe can be represented by means of a ‘Y’ piece of glass tubing with a balloon tied to each arm of the Y The limb of the Y is inserted through a rubber bung and the connection must be airtight to seal the interior of the bell jar from the atmosphere. The rib cage is represented by the bell jar, the rubber sheet represents the diaphragm. When the sheet is pulled down the pressure within the bell jar is reduced. Air is drawn in through the *Y’ piece into the balloons by atmospheric pressure and the balloons inflate. When the sheet is allowed to return to its raised position the balloons deflate. 2135 gineering.

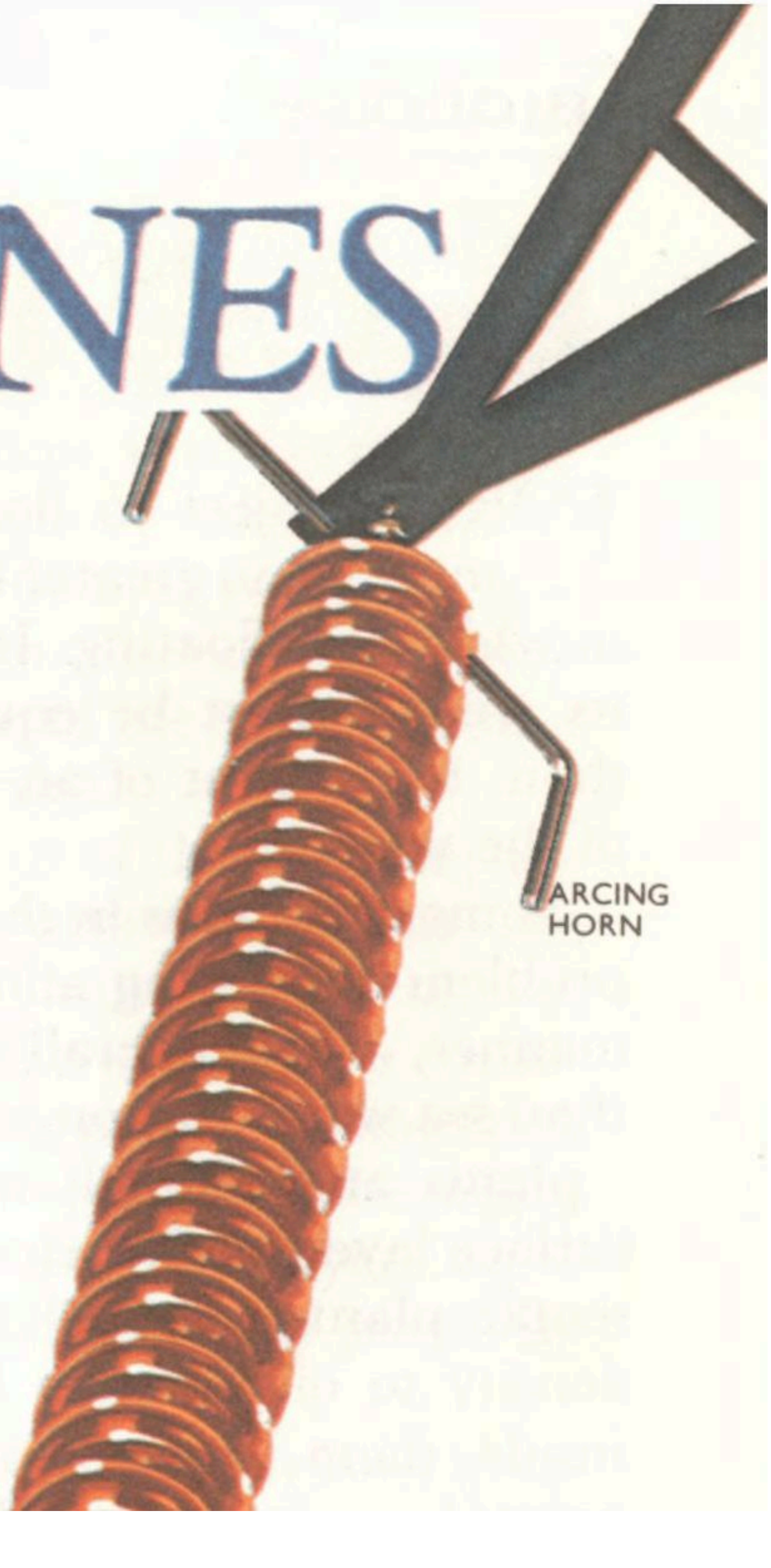


A model demonstrating the mechanism of breathing can be made using a bell jar with a rubber sheet pulled taut across its large open end. The lungs and windpipe can be represented by means of a 'Y' piece of glass tubing with means of a 'Y' piece of glass tubing with a balloon tied to each arm of the Y. The limb of the Y is inserted through a rubber bung and the connection must be airtight to seal the interior of the bell jar from the atmosphere. The rib cage is represented by the bell jar, the rubber sheet represents the diaphragm. When the sheet is pulled down the pressure within the bell jar is reduced. Air is drawn in through the 'Y' piece into the balloons by atmospheric pressure and the balloons inflate. When the sheet is allowed to return to its raised position the balloons deflate.

\ A i | Power Transport KES XxX N ESE XL ISAS iN x a HE beauty of a chain of pylons in the domestic electricity supply) and IN Ke iS carrying electric power across — most, in fact, use four. They must all Ke N \ the countryside is debatable, but it is be insulated from each other. ; san IN a plain fact that overhead power lines The wires carry Current, so they OS Vane T } are much cheaper than an under- become heated. Air cools the over\\ IN Sx + SS . | \ ground cable system. The overhead _ head power lines: extra water-cooling power line is a bare aluminium and equipment would be needed for underIN S \ \ steel wire. There isno need tosurround ground lines. y, WY it with an insulating plastic coating, For several reasons it is better to NN As oR % as in most domestic electricity cables. use four wires (or pairs of wires) per Pe * NIZE Air round the power line acts as an circuit, and the extra wires are easily ‘ at ex 2 = fj insulator, and additional insulation is | accommodated on a pylon. Most of = EE TSN necessary only at the points where the the large pylons in fact carry two \: line is supported by the pylon. An complete four-wire twin-conductor yj underground cable, on the other hand, circuits which share a common earthed needs insulation all along the line. wire. Three arms extending out from Digging the holes and laying the cable _ either side of the pylon support the is an expensive operation, and there _ wires, and space them so that the gap is possibility of corrosion. between individual wires is large It may be relatively cheap to trans- _ enough to insulate. The earth wire is port large volumes of gas, water and carried along the top of the pylon. oil underground, but large amounts Larger and stronger pylons could of electric power are quite a different carry three or even four circuits, so it proposition. Power has to be trans- would be impossible to use the same ported along the Grid and Supergrid system, slightly strengthened and ensystems at enormously high electric larged, to carry twice as much power pressure (or voltage, or electrical po- as at present. However, there are distential). The proportion of electric advantages. The pylon would become power lost through electrical resistance an unsightly monster, and it would be heating is less at high voltages than _ impossible to keep the circuits entirely at low voltages. But higher voltages separate. Grid circuits require mainnaturally require more effective insu- _ tenance, and when one circuit is being lation. This is comparatively easy to attended, the other circuits on the do on an overhead line, comparatively — same pylon must continue to transport difficult on an underground cable. power. If there are only two circuits Each power-carrying circuit needs __ per pylon chain, the maintenance men a minimum of three wires (correspond- _can work in safety. But it is dangerous ing to the earth, live and neutral leads —_ when there are more live circuits. Overhead or underground? The pylon carries The power lines of the grid form a an earth wire (at the top), and 6 twin con- _mprecnated ol transmission line. This is the name given ductors. Below: a single experimental — **P*R INSULATION L a underground conductor. to any system of wires carrying electricity from one place to another. The grid transmission line is designed to

- TRANSMISSION LINE A 275 kilovolt Supergrid conductor, steelcored and surrounded by aluminium wires. carry large amounts of electricity, and the first requirement is that it is a good electrical conductor. Silver is the best conductor, copper the next best, and aluminium comes third. Silver is obviously too expensive, aluminium is just over half as efficient as copper, but it has the advantage of being light. The lines look light and flimsy from the ground, but they set up enormous tensions in the pylons. Weight is of great importance, since a heavier wire sets up larger tensions and needs more pylons to support it. Aluminium wires can be supported by six pylons per mile: corresponding copper wires need ten per mile. This is even when the aluminium, a relatively weak metal, is wound around a heavier core of steel. Steel gives the transmission line mechanical strength and flexibility. Standard aluminium power conductors have a steel core of seven wires, surrounded by two or three layers of aluminium wires, consisting of 12, 18 and 24 wires. The aluminium also protects the more vulnerable steel from corrosion. Aluminium protects itself by forming a hard surface, oxide film. On average, six times per mile on the grid, and five times per mile on the Supergrid, a pylon has to be sited to support the line, keep the circuits spaced and well away from the ground. Pylons are usually made of steel, so it is impossible (as well as highly dangerous) to simply fasten the transmission lines to the steel supports. All the current would immediately be conducted down the pylon to earth. The supporting link must be an electrical insulator, capable of insulating voltages of up to halfa million volts. Since it bears the weight of the line, it must also be very strong. Two materials, high-grade porcelain and toughened glass, are commonly used. Up to 24 bell-shaped insulators fit into each other, pin to socket. The upper one is attached to the metal frame of the pylon, and the lower one has a supporting clamp for the transmission line. Two insulators, inclined at an angle, may be necessary when the transmission line changes direction and sometimes pairs of insulators are needed to support the twin conducting wires. At the upper and lower ends of each chain of insulators are sets of arcing horns, which prevent the line from being damaged when it is struck by lightning. Electric charge accumulates on the pointed ends of the horns, increasing the strength of the electric field in this region. The gap ‘between INSULATORS TWIN CONDUCTORS wspended® insulator, ee with arcing A typical 275 k carrying twin conducto horns. upper and lower horns is therefore the easiest current-route from conductor to earth. The gap is designed to hold back normal working voltages, but not the few million volts associated with a lightning discharge. A sudden over-voltage is given a quick route to earth, and the ‘flashover’ is kept clear of the transmission lines. The upper limit to the voltage used for overhead transmission lines is set by the leakage of charge away from the line—the corona discharge. This represents a power loss which increases as the voltage increases. 2137

LINES

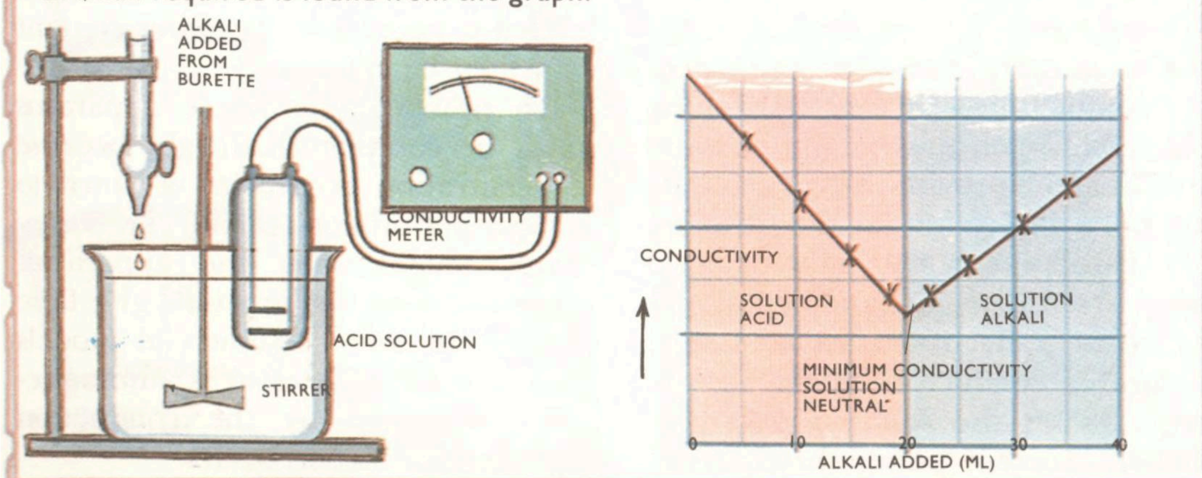


BIOLOGY KEEPING AFLOAT) R an object to float, its density must be no greater than the water in which it is floating. In other words, its weight must be equal to, or less than, the weight of an equal volume of the water. Some organisms in the sea solve the problem of keeping afloat in just this manner. Their overall density is less than sea water. Amongst the plankton ~—plants and animals drifting in the surface layers of the sea — some microscopic plants (diatoms) owe their low density to oil globules filling vesicles inside them. Some small protozoan animals are partially filled with water made less dense by saturation with DEPTH 5,000 FEET VOLUME OF SWIM BLADDER KEEPS THE SAME DEPTH — 15,000 FEET Some fish lower their density by inflating bladders within their bodies. Water pressure increases with depth; the effect is to compress the gas in the bladder. But more gas is secreted, keeping the fish at the same volume and hence at the same densit carbon dioxide or ammonia gas. Alternatively, to keep afloat a shape may be developed which has a very large surface area in comparison with the animal's total volume. The larger the surface area in contact with the water, the greater the frictional resistance to sinking. Small plants and animals living in the plankton may be flattened, ribbon-shaped or drawn out into long slender pencil shapes. Spines — projections from the surface of the organism — may be developed, both in plants and animals. As a rule, organisms in tropical seas are far more spinose than those in colder, denser waters. A bigger surface area is required to prevent sinking. Some organisms are actually able to modify the !ength of their spines according to the viscosity of the water. A third mechanism for keeping afloat is, of 2138 course, by swimming. Variations on a Theme Creating buoyancy by a lowering of density is found throughout the animal kingdom. But variations are found both in techniques and structures used. Some of the most highly evolved of the jellyfish family (the coelenterates), despite their large size and overall weight, have solved the problem by the development of air sacs. Air sacs increase the volume of the body without contributing much to an increase in weight. The result is a lowering of density. Perhaps the best known is Physalia— the 'Portuguese man-ofwar' which has an enormous bulbous air sac (pneumatophore) — probably developed from a fold of skin. The sac floats above the surface of the waves and supports the rest of the animal. A similar device is used by the Sargasso seaweed which keeps afloat using air-filled bladders. Janthina, a planktonic snail, actually makes its own bubble raft. Gas is secreted and contained in thin films of mucus. For animals which, in their search for food, move from one level to another, the ability to both rise and sink without spending energy in swimming is a great advantage. Such an ability requires a method of controlling the animal's density — that is, its weight relative to its volume. Some fish possess a swim bladder — a balloon-like sac positioned above the gut. Gas is secreted into the bladder from the blood system. The weight of the heavy bony skeleton and muscle tissues which cause the fish to sink is off-set by the lightness of the inflated bladder. The result is that the fish can float without expending energy —a freedom which no doubt has been a great evolutionary advantage. At lower depths, however, the pressure of the water increases, the gas inside the bladder is compressed. The volume of the fish is lowered, its density is increased. Counteracting this effect, more gas is absorbed by the bladder from the blood, so keeping the sac at a constant RAFT OF BUBBLES IANTHINA A PELAGIC SNAIL volume. The gas pressure may rise inside the sac to an enormous degree. For instance, fish swimming at 15,000 feet need a pressure of 7,000 pounds per square inch inside the bladder to withstand the sea's pressure. Not all fish have swim-bladders. The sharks and their allies (the cartilaginous fish) have never had lungs and consequently have never developed swim-bladders. They constantly swim in order to keep up, and are not nearly so stable. Some degree of buoyancy is, however, obtained from special oils in the liver. Of the bony fish, hottom-livers such as the plaice have lost their swimbladders altogether. So have some fast-swimming bony fish — the mackerel and the tunny. The reason underlines one of the drawbacks of the swim-bladder as a hydrostatic organ. In changing its volume according to the depth of the sea, some length of time is required. Secretion and abEarly Nautiloids probably secreted air into the chambers of their shells just as Nautilus does today.

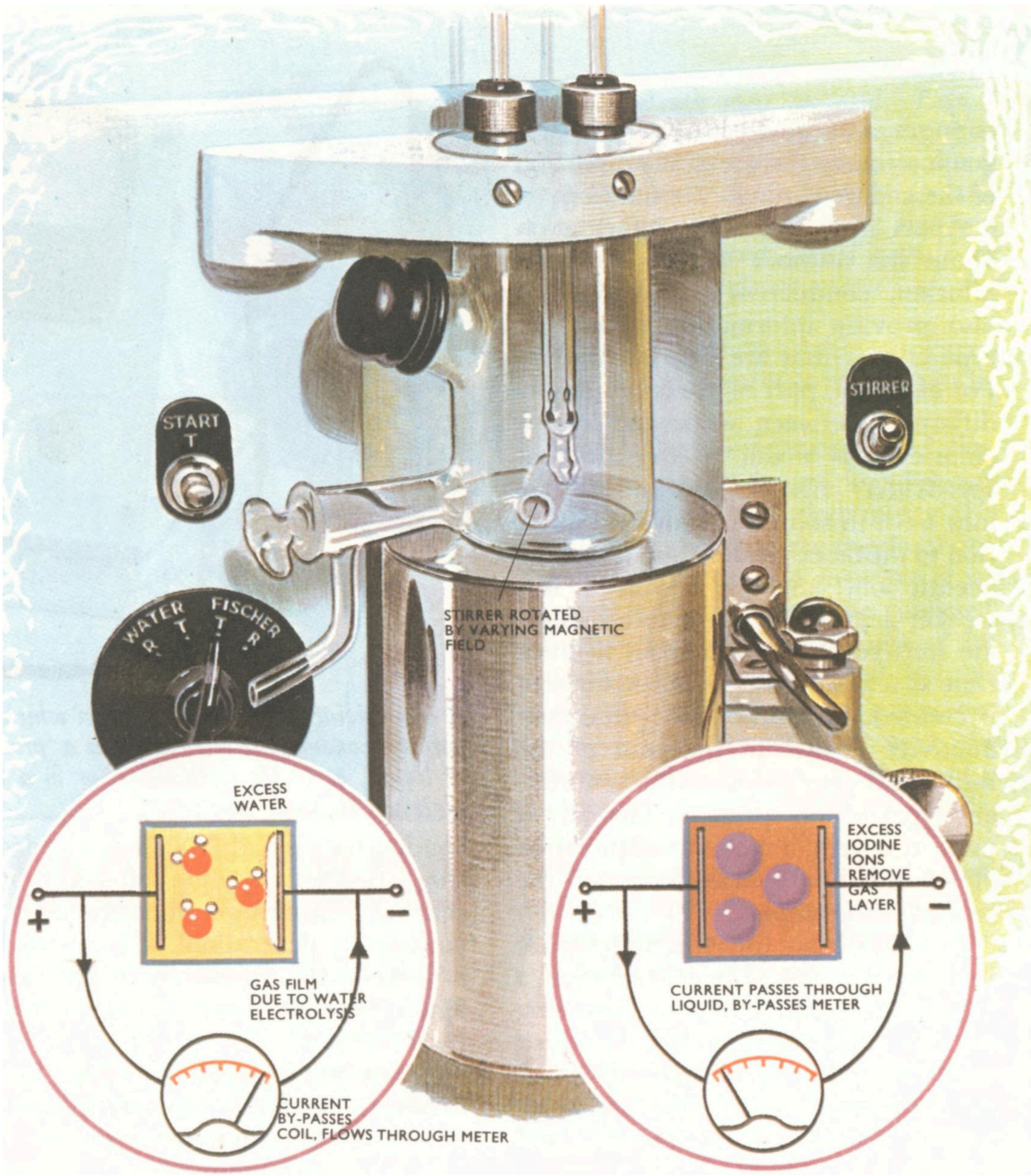
MODIFIED APPENDAGES Microscopic organisms demonstrate ways of keeping afloat; if you cannot float you must swim (right). Extreme left: lanthina, supported by a raft of air bubbles ; extreme right: Physalia, an advanced member of the jelly-fish family, supported by an azr sac. sorption of air is a delicate operation. Thus if a fish caught in a trawl net is brought rapidly to the surface, the bladder will burst, causing damage to the other internal organs. Consequently mackerel and tunny, though they must swim to keep afloat, nevertheless may move rapidly from one level of the ocean to another. The volume of a solid is not so influenced by high pressures as gases, and in some fish a strong development of fatty tissues is found around the lung. But fat has a density of 0-9 — only slightly less than water, and consequently must be present in fairly large quantities to be effective. The cuttlebone of cuttlefish not only gives support. Spaces within the 'bone' (inset) can be filled either with air or water so changing the creature's density. SPACES CAN BE FILLED WITH AS OR WATER © PHYSALIA ~ THE PORTUGUESE "MAN O' WAR One way of lowering density is to lose weight compared with volume. Some deep-sea fishes have their bones and swimming muscles and the rear parts of their skeletons so reduced that they can hardly swim. The front parts of the body are not affected and the fish lie in wait for prey. One other mechanism of lowering density is to dispense with as much as possible of heavy tissues. Thus some bottom-dwelling ocean fish have a much reduced skeleton in the tail region. The swimming muscles are poorly developed and appear delicate, almost transparent. The proportion of dilute body fluids and fatty tissues is increased and the body density is just less than water. Among the invertebrate marine creatures the cephalopods —a group of molluscs—are the most highly evolved. They live in shallow waters about the coast as well as in the deeps. A very spectacular cephalopod is Nautilus — the last survivor of a stock » that goes back 500 million years. The » colourful spiral shell in which it lives ' is divided into chambers. The outside _ chamber is the largest and it is here that the animal is situated. The other AIR SAC OR PNEUMATOPHORE chambers have become gas-filled. A strip of tissue extends back through the chambers and appears responsible for the gas secretion. It is this development of a buoyancy mechanism in early cephalopods that probably first enabled them to develop an active swimming mode of life rather than a bottom-living habit. Most numerous of the cephalopods today are the cuttle-fish. Along the back of this creature is the chalky internal shell — the well-known cuttlebone. The cuttlebone, when examined in detail, is found to be a beautiful structure consisting of layer upon layer of calcite plates neatly joined by calcite columns. The rigidity of the cuttlebone gives the cuttlefish support as it swims through the water. But also the bone gives buoyancy for, being largely hollow, its density is 0-6. The cuttlebone is not just a static organ, its density can change. When the cuttle~ fish wishes to sink, the chambers in ‘ the cuttlebone are flooded with water. When rising, the water must be — pumped out again —a process which™ appears to be carried out by osmosis: Yet another variation is found ine some deep water squids — allies of the: cuttlefishes. These squids have solid chitinous internal shells which, though acting as a skeleton, have no hydrostatic function. Nor is there any kind of air-filled bladder present, or much fatty tissue. j The riddle of the squid’s floating — \ ability — for it can float at all levels — rests in its body fluid. The body fluid occupies two-thirds the weight of the squid but has a density of about 1-01 compared with sea water’s 1-02. The density of the fluid is low because of ammonium ions present in solution — ammonia being the nitrogenous waste. 2139

LAMP FILAM BRIGHT — lon Measurements and Titrations HEN liquids conduct electricity there is a two-way traffic in ions. Positive ions move to the negatively charged cathode and negative ions to the positively charged anode. At the electrodes the ions lose their charges to form neutral atoms or molecules. So the ions, which were originally in solution, are released in the form of gases or layers of plated-out metal on the electrodes. This process of electrolysis has many practical applications — the important electroplating industry is based on it, and the industrial separation of many chemicals — chlorine gas and _ aluminium amongst them -—is carried out in large electrolytic cells. To the scientist, the study of the properties of conducting solutions yields important information. An easily-measured quantity is the electrical conductivity of a solution — this 1s a measure of the amount of electric charge that is pushed through the liquid (as a result of the two-way ion traffic) when a voltage difference 1s set up between the electrodes. The conductivity values of different 2140 ENT MODERATELY e MODERATE CURRENT SOURCE Of ALTERNAT| CURRENT ASOLUTION OF ACETIC ACID LAMP FILAMENT DIMLY LIT solutions are of interest to the chemist because they give a direct guide to the numbers and types of ions present in a solution, and this, in turn, provides information about the chemical nature of the compound dissolved in the solution. This is easily demonstrated in a simple experiment. Solutions of hydrochloric, sulphuric and acetic acids are made up so that each has the same normality (this means that equal volumes of each solution would be neutralized by equal amounts of an alkali such as sodium hydroxide). Equal volumes of each acid solution are poured into identical conductivity cells. These are glass cells in which identical pairs of metal electrodes have been sealed. The three cells are each placed in series with an identical electric light bulb and the three pairs of bulb and cell are all connected in parallel across the electricity mains. An alternating current passes through each cell and bulb, and the three bulbs light up. The lamp in series with the hydrochloric acid cell shines brightest of all, and the 'sulphuric acid' bulb is brighter than the bulb in series with the acetic acid. This can only mean that the current passing through each of the three cells is different, so the In the three cells there are acids of equal concentration but different strengths. The cells are connected across a common mains supply but the three bulbs light up with different intensities. This shows that each cell carries a different current because the solutions contain different numbers of ions. number of ions available to carry \charge to the electrodes is different in each cell. What is the significance of this ? The number of free hydrogen ions present in an acid solution determines the strength of an acid, so the conductivity of an acid solution is a measure of its strength. (This is not to be confused with the normality or concentration of the solution — this is determined by the total number of acid molecules present.) The strength of an acid will determine the vigour of its reactions — the rate at which it will attack metals, for example. When an acid is dissolved in water there is immediate dissociation into ions. Hydrochloric acid molecules ionize into hydrogen (H"*) ions and chlovide (Cl) ions: HCl > H*+C!. The extent to which dissociation takes place determines the strength of the acid. In a strong acid such as hydrochloric acid 92% of the acid molecules dissociate into ions. In acetic acid, a weak acid, only 1% of the acid molecules dissociate. The electrical conductivity is obviously related to the strength of the acid because the hydrogen ions carry the greatest proportion of the charge through the solution. Apart from providing information about the number of free ions in a solution, conductivity measurements also provide information about the type of ions that are free in solution For example, part of the conductivity difference between acetic and hydrochloric acids is due to the different speeds that chloride ions and acetate ions have in solution. This is in addition to the fact that there are less of the acetate ions, anyway, because it is a weaker acid. More of the faster chloride ions reach the anode in a given time so a greater negative ion current is created. In solution, the total conductivity is due to the sum of the individual conductivities of the different types of ions present. This fact is used in a very important application of conductivity measurements — the conducti metric titration. In an ordinary titration, to measure the concentration of an acid solution, a measured quantity of a neutralizing solution (alkali in this case) is added until the end point is reached. Then the hydrogen (acid) ions are completely neutralised by the hydroxyl (alkali) ions: BoP OF », 1,0 and a chemical indicator like litmus or phenolpthalein will change colour. In a conductimetric titration the conductivity of the solution is continuously measured. At first, when the solution is strongly acid, there are many hydrogen ions present and the ~~ PORTABLE CONDUCTIVITY PROBE A commercial conductivity motor in which a direct reading of conductivity is obtained. The portable conductivity cell is used as a 'probe' which may be inserted into a stationary volume of liquid, as in the illustration, or in a flowing liquid. Inside the cell a pair of parallel platinum electrodes are sealed. conductivity is high. As alkali is added, the conductivity is lowered and reaches a minimum when the acid is completely neutralised. If more alkali is added, the conductivity increases again because there is an excess of hydroxy] ions. Quantities of alkali are added from a burette and the solution is mixed before the conductivity is measured. The measured conductivity is plotted on a graph, against volume of alkali added, and the curve reaches a minimum for the volume of alkali needed for neutralisation. Ideally, the numbers of free hydroxyl and hydrogen ions present is then very small indeed. The method provides an accurate way of determinThe principle of the conductimetric titration. The acid in the beaker is neutralized by alkali that is added, drop by drop, from the burette. The liquid is continuously stirred and the conductivity is measured. When the solution is strongly acid its conductivity is high but this decreases as alkali is added until the neutral point is reached. When too much alkali is added the conductivity rises again. At the minimum, the exact amount of alkali required is found from the graph. ALKALI ADDED FROM BURETTE ing end-points, and it is used in a number of titration reactions where chemical indicators are not satisfactory. Measuring Conductivity The solution is poured into a glass conductivity cell in which a parallel pair of platinum electrodes is sealed. The conductivity (K) ofa solution between plates of area (A) separated by distance (D) is given by K = rae where R_ is the measured electrical resistance between the plates. It is not always practicable to find the conductivity of a solution from the dimensions of the cell and its measured resistance. Then, the resistance is compared with that obtained when the cell is filled with solution of known conductivity. The cell, filled with electrolyte is placed: in a bridge circuit and its resistance is compared with that of a known resistor. In the bridge circuit, an alternating current is used so that gas is not given off at the electrodes — it would affect the measured conductivity. In the Kohlrausch bridge circuit the value of resistance of the cell is ‘matched’ with that of the circuit resistance, using the bridge wire. (This is very much like using the Wheatstone bridge to measure a d.c. resistance.) When balance is achieved no sound is heard in a pair of earphones connected to the circuit. 2141

The principle of the conductimetric titration. The acid in the beaker is neutralized by alkali that is added, drop by drop, from the burette. The liquid is continuously stirred and the conductivity is measured. When the solution is strongly acid its conductivity is high but this decreases as alkali is added until the neutral point is reached. When too much alkali is added the conductivity rises again. At the minimum, the exact amount of alkali required is found from the graph.



Applications . Conductimetric measurements are carried out in industry to determine the purity of water. Distilled water has a low conductivity (if really pure there is a low ionic concentration), so to test purity, conductivity meters are employed. The degree of purity of inflowing feed waters for' industrial boilers must also be controlled to prevent scaling. Conductivity meters are used in this work as well. Specially constructed probe cells are available. Finding Moisture Contents One interesting application of conductimetric titration is for measuring small moisture contents in liquid and solid materials. The method used is called a Karl Fischer titration and it serves as an example of electrometric titration where a simple acid-alkali reaction is not monitored. The sample whose water content is to be measured is mixed with methanol so that the water is dissolved out. The methanol-water mixture is then titrated against a solution of Karl Fischer reagent. This is a solution in methanol of pyridine, sulphur dioxide and iodine. The chemical reaction that occurs when the water is mixed with the reagent is a complicated one but the overall reaction may be written: H,O + CH;0OH + SO, + I, = 2HI + CH;SO,H If insufficient reagent is added free water is present in the solution. When the end-point is passed there are free iodine ions present. The problem is to find this end-point. When an excess of iodine is present the solution changes colour from a yellow to deep brown but the endpoint cannot be detected visually with an accuracy sufficient for measurement of very small quantities of water. In Karl Fischer titration this is done automatically by detecting a change in conduction in the reagent beaker. A pair of parallel platinum electrode wires is placed into the beaker and a steady voltage difference maintained' between them. When there is an excess of water present the electrodes become polarized. This means that a film of the gas produced by ¢lectrolysis in the water adheres to electrode surfaces. The result of this is to 'break the 2142 EXCESS WATER COIL, EuOWS THROUGH METER CURRENT PASSES THROUGH LIQUID, BY-PASSES METER Part of an automatic Karl Fischer titration apparatus, to which the reagent 1s added automatically. (Inset) When water is present a gas film 1s formed on the electrode and all the current passes through the meter. When there 1s an excess of iodine the film disappears and the reagent 'shunts' the meter producing zero reading. Volume of reagent added to produce zero reading gives an accurate value of moisture content. circuit' between the electrodes, preventing current passing through the liquid from one electrode to the other. When there is an excess of iodine present, due to addition of more Karl Fischer reagent, the iodine ions have the effect of depolarizing the electrodes, breaking the film so that a current passes between the electrodes. The change from ‘no conduction’ to ‘conduction’ is very sensitively dependent upon the amount of water present and it provides a very accurate end-point indication for the titration. The end-point is normally detected using a sensitive galvanometer, and the titration electrodes act as a shunt (a parallel conductor) to the instrument. When the solution does not conduct, current passes through the galvanometer so there is a deflection reading on the dial of the instrument. When the solution conducts, the current passes through it, bypassing the. meter so there is no deflection. The Karl Fischer reagent is then titrated against the water solution until there is no deflection. The volume of reagent added is noted and the water content can then be calculated. In practice, the whole apparatus must be completely airtight because concentration accuracies of down to 5 parts per million (p.p.m.) are sometimes sought with the apparatus. Moisture from the air would give false results. In some commercial models the addition of reagent is automatically controlled by the conduction signals from the electrodes.


STIRRER
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| MATHEMATICS Significant Figures NE of the nearest stars in the sky is some 207,000,000,000 miles away from the Earth. This does not mean that the star is exactly 207 thousand million miles away. Its distance is being quoted as a round figure, to the nearest thousand million miles. It would certainly be unjustified to state that the star was 207,402,410,711 miles from the Earth, because measurements of stellar distances can never be made to the nearest mile. In measuring vast distances, the error can be quite considerable—a few hundred million miles either way. Another factor makes nonsense of quoting too many figures. As the Earth orbits the Sun, its distance from a star can change by 186,000,000 miles, the approximate diameter of the Earth's orbit. So it is meaningless to give the star's distance to even the nearest million miles, unless the position of the Earth at the time is quoted as well. To show that measurements are inexact, or that the distance fluctuates, the noughts are inserted instead of any other numbers, and the distance is given to the nearest round figure. Bathroom scales cannot measure to the nearest gram, and the weight fluctuates with the time of day. Only 2 significant figures can be quoted. <2. gseuae WEIGHT AFTER MEAL - 6! KILOS way oe eS tl pt WEIGHT AFTER POCKETS ARE EMPTIED -59 KILOS Lo A 4 in a number always stands for four, and a 7 always stands for seven, but o does not necessarily stand for the number nought. In this example, the noughts after the 7 are fill-ins, and they show the order of magnitude of the distance. But the o between the 2 and the 7 obviously does stand for the number nought. It is said to be a_ significant figure. The other noughts are not significant. The significant figures in the number 0-00067 are 6 and 7. The noughts are not significant, showing only that the 6 means six-ten-thousandths, and the 7 means seven-hundred-thousandths. The significant figures of any number are the ones which are known beyond any shadow of doubt. They are the figures which should be quoted in the results of any experiment. If length measurements are involved they should not be quoted to the nearest millimetre when the accuracy of the measuring instrument is only to the nearest centimetre. If, during the calculation stage, inaccurately-known values are multiplied and divided by each other, it is possible to end witha long string of figures by continuing the long multiplication or division longer With care, and a good | metre rule, it should be possible to measure to the nearest centi- | metre, i.e. to 4 significant figures— 12-32 _ metres. cab The Index Notation If all the

noughts after 207 in 207 ,000,000,000 are insignificant, there is little point in writing them down laboriously each time the number | appears in a calculation. It is better to use a shortened form of notation, the index notation. 207 ,000,000,000 becomes 207 x 10°. 10° is another way of writing a thousand million, i.e. a | followed by 9 noughts. 9is called an index number. 207 x ,10° is also equal to 20-7 X 10'° and 2-07 x 10"'. The last of these is best for showing the order of magnitude of the distance. / The distance is of the order of a hundred thousand million miles (10" miles or | |-followed-by-|1-noughts miles). Another way of writing 0-00067 is 6:7 x 10~*. 10~' is the index notation for 7; 1072 is the notation for 7; 10-4 is the notation for 950; and 0-00067 is just the same asza-o00.Besides reducing the number of [| figures needed to write down a number, the index notation makes multiplication and division easier. than is necessary. Only the first few figures are significant. The number of them depends on the accuracy of the measurements. The answer should be corrected 'to the nearest significant figure'. For example, if the answer comes to 1812, and the figure 2 cannot be justified, it is written as 1810. But if the answer comes to 1817 at the end of a calculation, its value, to the nearest significant figure, would be 1820. eta? ie a eg hg er Many more exact m easuring instruments could be used, so that the length could be quoted to the nearest millimetre (i.e. 5 significant figures) and so on. But there is little point in measuring more accurately because the paper could stretch or shrink. 2143

The Index Notation

If all the noughts after 207 in 207,000,000,000 are insignificant, there is little point in writing them down laboriously each time the number appears in a calculation. It is better to use a shortened form of notation, the *index notation*.

$$207{,}000{,}000{,}000$$

becomes

$$207 \times 10^9.$$

10^9 is another way of writing a thousand million, i.e. a 1 followed by 9 noughts. 9 is called an *index number*.

207×10^9 is also equal to
$20 \cdot 7 \times 10^{10}$ and $2 \cdot 07 \times 10^{11}$.

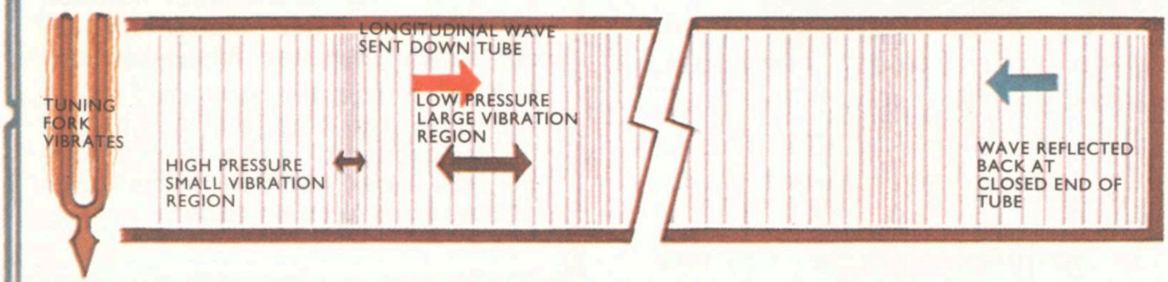
The last of these is best for showing the order of magnitude of the distance. The distance is *of the order of* a hundred thousand million miles (10^{11} miles or 1-followed-by-11-noughts miles).

Another way of writing $0 \cdot 00067$ is $6 \cdot 7 \times 10^{-4}$. 10^{-1} is the index notation for $\frac{1}{10}$; 10^{-2} is the notation for $\frac{1}{100}$; 10^{-4} is the notation for $\frac{1}{10{,}000}$; and $0 \cdot 00067$ is just the same as $\frac{6 \cdot 7}{\ldots}$

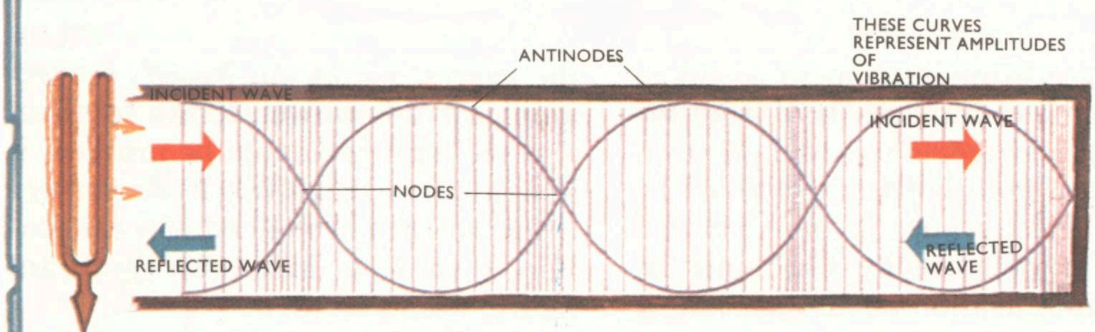
0·00067 is just the same as $\frac{\quad}{10,000}$.

Besides reducing the number of figures needed to write down a number, the index notation makes multiplication and division easier.

Seein OUND is a form of Wave motion, carried from one place to another by vibration of molecules in solids, liquids or gases. When a tuning fork is struck, the prongs are set into motion, back and forth, and the prongs set up vibrations in the surrounding air molecules. The air molecules follow a form of motion very similar to the prong itself — as the prong moves outwards a thin layer of air molecules is pushed outwards with it. The molecules in the layer then not only return to their original | ENeNT POWNITUBE | |) LOwiPRessure | | l “LARG VIBRATION _ ||| ube yenArIN HIGH PRESSURE Ott Ph SMALL VIBRATION | | LH REGION Stationar i j ' } ! This is a form of longitudinal wave * the vibrations occur in a direction along the direction of motion of travel of the sound. The rate at which vibration ‘bundles’ are sent out from the fork depends on the frequency of vibration of the prongs — this determines the pitch of the sound that is heard. The distance between each vibration bundle is the wavelength of the sound. Sound is emitted in this form, as a moving pattern of vibration waves. But in many musical instruments — an ' Phi it] Phiyd | itt WAVE REFLECTED || | Babses eu be! ' OF |‘Fbe Pitti Above: The tuning fork sends layers of air molecules into vibration. The longitudinal waves pass up the tube. Layers performing large vibrations follow, layers of small vibration. It is regions of vibration that move along the tube, not air molecules themselves. At the closed tube-end the vibration waves are reflected back. Below: Ina closed tube a stationary wave is created by the ‘mixing’ of the incident and reflected waves. In the standing wave pattern produced,nodes always occur at the closed end, antinodes at the open end. ANTINODES positions, but move beyond (like a pendulum bob). But before they do so, they push out the molecules in the adjoining air layer. These molecules then swing out and return to their original positions and through the other side and back again. In the process, they set the next layer into vibration. In this way, the vibration of the prong of the fork is copied by vibrations in successive layers of air. The vibrations travel outwards from the fork. It should be remembered that it is only the vibrations of air that are being sent out from the fork; there is no transfer of the air itself. 2144 THESE CURVES peer AMPLITUDES organ pipe, for example, there are stationary patterns of waves set up. An organ pipe is stimulated into giving out a musical note when air is blown into its end. A regular pattern of vibrating layers of air molecules is set up in the pipe. y W aves In the pipe, a region of vibrating layers of molecules merges into a region of no vibration. This merges into another region of vibrating molecules. The regions of maximum vibration are called anti nodes, those of no vibration, nodes. In a closed pipe it is obvious that a node must exist at the closed end —the molecules cannot possibly vibrate against a fixed surface. The antinode exists at the open end. The stationary wave (or standing waue pattern) is formed when a sound wave travels down the tube and is reflected at the closed end. The two waves, forward (incident) and reflected, 'mix' together and interfere. At the nodes, the two waves always cancel each other out, but at the antinodes the waves add to each other, so the nodes represent a positions of no vibration, but the antinodes represent a position of maximum vibration, of air molecules. The existence of nodes and antinodes can be demonstrated in a number of different ways. The best-known demonstration is in Aundt's tube. A long clamped rod is attached to a flat plate within a glass tube, which is set into vibration when the rod is stroked with a resined cloth. The plate sends out sound vibrations into a long glass tube. Into the tube is sprinkled fine lycopodium powder or sand. The rod is moved into the tube so that the sound waves it gives out 'fit' the tube — an antinode is at the open end when there is the node at the closed end. (The tube is then said to resonate.) At antinodes the vibration of the air 'sweeps' the powder away and it accumulates at positions of no vibration — the nodes. The resulting regularly-arranged piles of powder along the tube represent the position of the nodes. In Kundt's tube the positions of the nodes are shown where the small piles of lycopodium powder collect. Stroking the rod makes the attached plate vibrate and the standing wave pattern is produced in the tube. The distances between the nodes 1s one-half of a wavelength of the sound. ROD STROKED WITH ROSINED CLOTH SETS UP VIBRATIONS TUBE CAN BE FILLED WITH SELECTED GAS, USING THESE INLETS _ HEAPS OF POWDER | } FORMED AT NODES

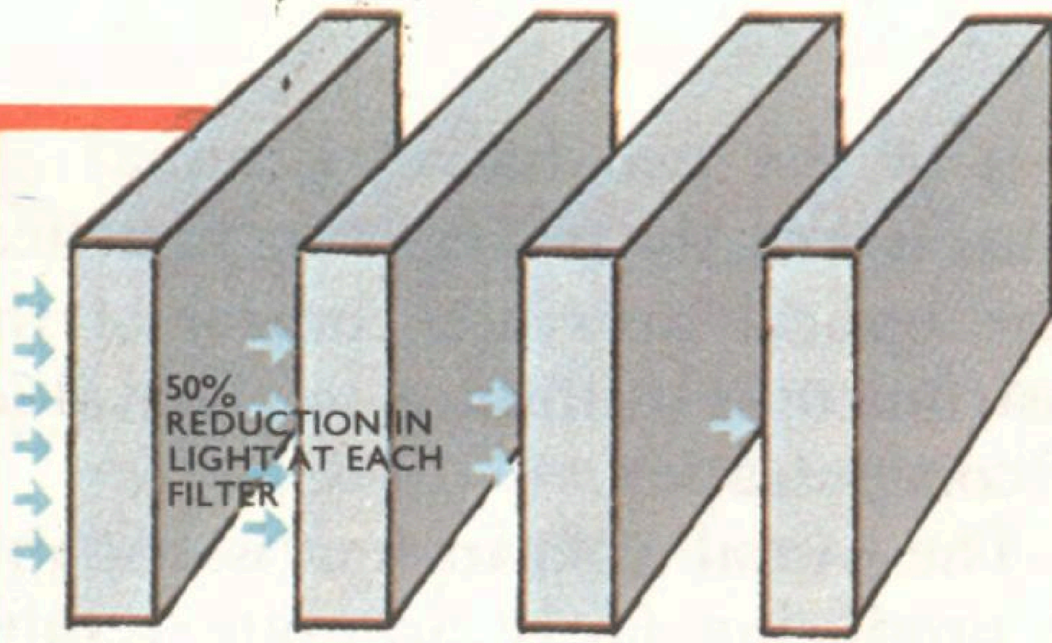


Above: The tuning fork sends layers of air molecules into vibration. The longitudinal waves pass up the tube. Layers performing large vibrations follow, layers of small vibration. It is regions of vibration that move along the tube, not air molecules themselves. At the closed tube-end the vibration waves are reflected back. Below: In a closed tube a *stationary* wave is created by the 'mixing' of the incident and reflected waves. In the standing wave pattern produced, nodes always occur at the closed end, antinodes at the open end.

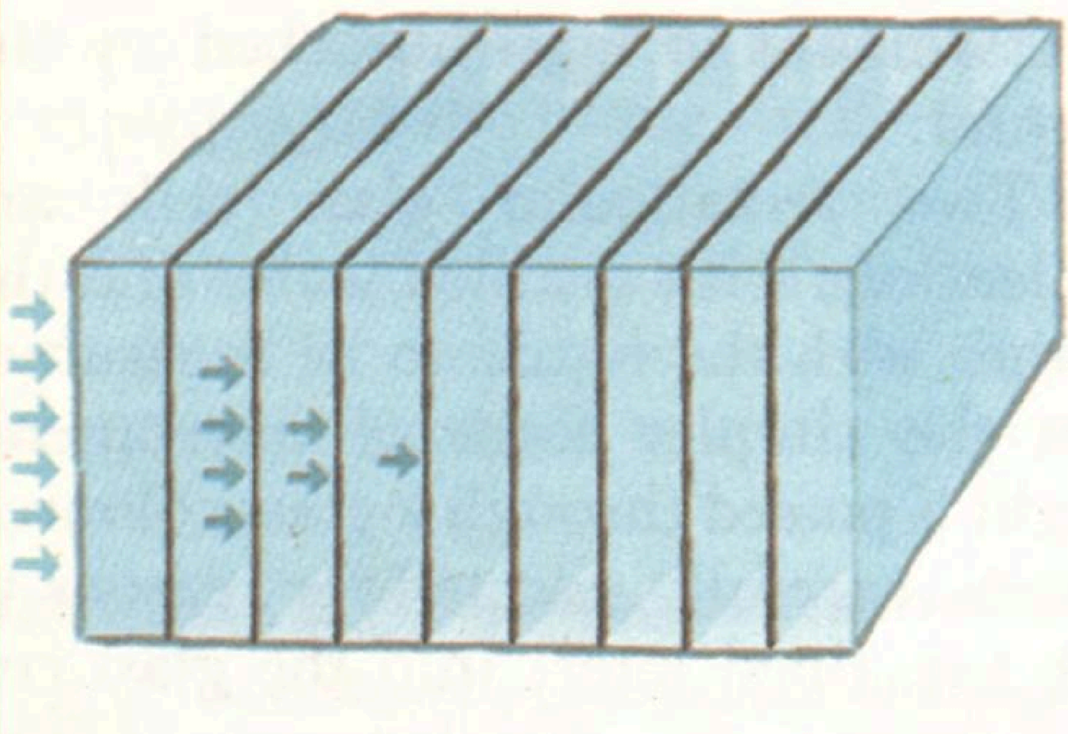


A simple commercial colorimeter. The instrument 1s set to read 'zero' by opening the light aperture with a 'blank' specimen (of water) in the test tube. The test tube is then filled with the liquid whose concentration Optical LIGHT SCIENTIFIC INSTRUMENTS is to be determined and its density 1s read off the meter dial. f COLOUR FILTER PHOTOCELL Density Measurements IN industry and the research laboratory the scientific instrument designer and manufacturer plays a vital role. Very often, an instrument is developed that measures some quite simple property of a material. But, if the instrument is designed to make accurate and selective measurements, the results can be of great value to the scientist. A good example of this is in the measurement of the amount of light that is absorbed by a specimen of a material (gas, liquid or solid). The quantity measured is normally the optical density. This measurement gives an indication of the proportion of light that is cut out by the specimen. The accurate measurement of optical densities can provide valuable data for research workers in fundamental science, the analytical chemist, as well as the photographic technician. Optical Measurements in Chemistry Some of the most useful density measurements that are made in chemistry are in colorimetric analysis. The use of this method depends on Beer's Law. This states that the light absorbed by a coloured solution depends on the number of molecules of the coloured compound dissolved. Since the number of molecules dissolved is, in fact, the concentration of the solution, the measurement of the light absorbed gives direct indication of the concentration of the solution. This has given rise to the technique of colorimetry in chemistry and such absorption measurements are made using colorimeters or spectrophotometers. In the Pulfrick instrument, the light is passed through a cell containing a specimen of the solution of known concentration. This is matched with light passing through a cell of the same solution, but of unknown concentration. Matching is achieved by varying the light from the 'unknown' cell by altering the opening of an optical aperture. As the aperture is opened or closed, the light is varied. In the optical system of the instrument, light from a common source is split so that it passes through the two cells. The eye then sees a circular field that is split into two halves, and when light from the 'unknown' cell is adjusted to the correct value the brightness of the two halves is equal —in other words, the two cells are matched. The optical aperture control of the instrument is calibrated to give values of concentration. The visual colorimeter is capable of producing fairly accurate results, but has now been superseded by the photo-electric colorimeter. In this instrument, the human eye is replaced by a photo-electric cell, which measures the amount of light absorbed by the liquid. Two rectangular glass cells are filled, one with distilled water and the other with the liquid to be measured. In the simpler form of instrument, light is passed through a glass coloured filter (carefully selected for reasons to be explained later) into the glass cell containing the distilled water. The light is collected by a photocell which produces a current that is displayed on the dial of a sensitive galvanometer. The meter reading is adjusted so that a full-scale deflection, at the '100% transparency' or '0-00 density' mark, is indicated for the distilled water. The water cell is then replaced by the cell containing the 'unknown' solution, and a second reading taken. The amount of light received by the photocell is now reduced by absorption in the cell, so the meter deflection is smaller. The reading on the calibrated IN ALLOY, THERE ARE FEWER TOMS OF ELEMENT. PHOTOCELL SMALLER DENSITY READING FOR LINE A THAN FOR LINE B Line density measurements are used in spectroscopy to determine the concentrations of the elements in, for example, an alloy. There are more atoms of element B so density of the line B is greater. 2145

$e tse If each filter cuts down light by 50%, two filters will cut it down to 25%, three to 124%, four to 64%. Lbbtss When light is passed through a column of coloured liquid, some of the light is absorbed. The amount absorbed is not proportional to the length of the column — doubling the length does not result in halving the amount of light coming out. In fact, the light decreases with column length according to an exponential law. One way of expressing this law is to say that if light of intensity I, enters the solution and light of intensity | comes out, then the logarithm of + is proportional to the length of the column. The quantity log 2 is called the optical density, and this definition is used to specify the lightreducing property of any material that absorbs light. Not only is the optical density Proportional to the length of the column, but, according to Beer's law, to the concentration of a coloured solution. For this reason a colorimeter measures the optical density, and the solution concentration is found from values of optical density. dial gives a direct indication of optical density of the liquid. Selecting the Wavelength Band To obtain the most accurate results with any colorimetric method it is essential that the wavelength of the light be carefully selected. A red solution appears red because it absorbs light from parts of the visible spectrum other than red and allows the red light through. If white light were passed through a solution even of high 2146 concentration a great amount of light would still be transmitted. Small changes of concentration in the solution would not have much effect on the amount of light absorbed. This would mean that the sensitivity of the instrument — the change in photoelectric current with change in solution concentration — would be very small. For this reason, light of the colour that is absorbed most strongly by the particular solution is passed through the cell. Then a small change in concentration leads to a big change in absorption, and to a very sensitive measurement of concentration. In the red solution, blue light is therefore passed through because it is absorbed strongly. In a blue solution red light is passed through. The light of different colours is selected by choice of appropriate colour filters, placed in the light path of the glass cell. In more accurate instruments, the wavelength of the light used must be even more carefully selected. In the spectrophotometer a very narrow light 'waveband' is selected, produced by a monochromator. The monochromator is essentially a prism spectroscope in which light of the required wavelength is selected by rotating the prism. The light refracted in the instrument falls against a narrow slit aperture which only allows through light in a narrow wavelength range. Concentration measurements can be made to a high degree of accuracy with such instruments, which in practice are made with a number of different refineAn accurate spectrophotometer. An image of the light source 1s directed to the collimating mirror and then on to the prism mirror combination. The wavelength of light used is controlled by rotating the prism and is returned along its original path and then through the absorption cell. It then passes on to the photocell. ments. There are, for example, special circuits designed to ensure accurate measurement of photocell current. But another important use of the spectrophotometer is in measuring absorption over a complete range of wavelengths rather than at a single wavelength. The specimen can be either solid or gas, and by 'scanning' the absorption over the complete range, valuable analytical information can be obtained. Different chemical groupings and types of bonds have their own characteristic absorption 'bands', and by studying the absorption spectrum, their presence can be detected. There are instruments available that automatically scan complete spectral ranges and provide charts displaying (Left) The absorption curve of potassium permanganate solution.. Light is most strongly absorbed at the wavelengths corresponding to the peak of the curve. Density measurements are best made with light in this wavelength range — concentration can then be found using the optical density-concentration curve (right). WAVELENGTH OF LIGHT OPTICAL DENSITY

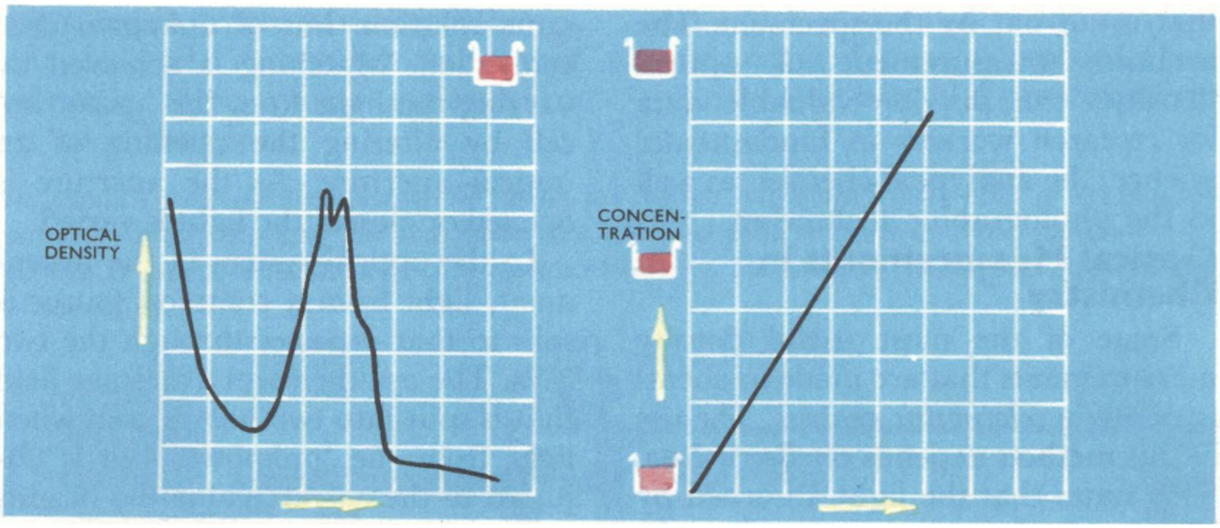


If each filter cuts down light by 50%, two filters will cut it down to 25%, three to $12\frac{1}{2}$%, four to $6\frac{1}{4}$%.

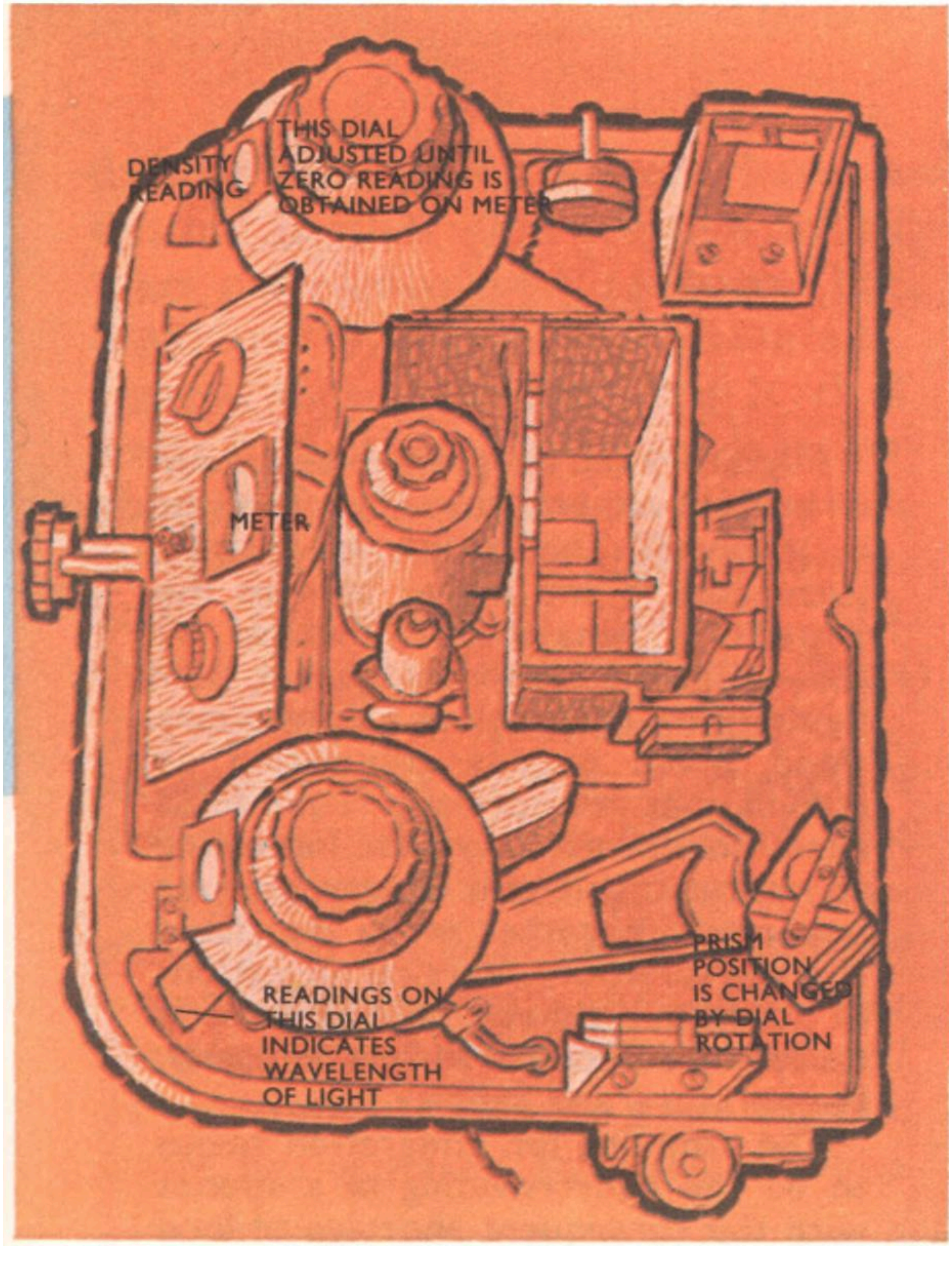


When light is passed through a column of coloured liquid, some of the light is absorbed. The amount absorbed is not proportional to the length of the column – doubling the length does not result in halving the amount of light coming out. In fact, the light decreases with column length according to an *exponential law*. One way of expressing this law is to say that if light of intensity I_0 enters the solution and light of intensity I comes out, then the logarithm of $\frac{I_0}{I}$ is proportional to the length of the column. The quantity $\log \frac{I_0}{I}$ is called the *optical density*, and this definition is used to specify the light-reducing property of any material that absorbs light.

Not only is the optical density proportional to the length of the column, but, according to Beer's law, to the concentration of a coloured solution. For this reason a colorimeter measures the optical density, and the solution concentration is found from values of optical density.

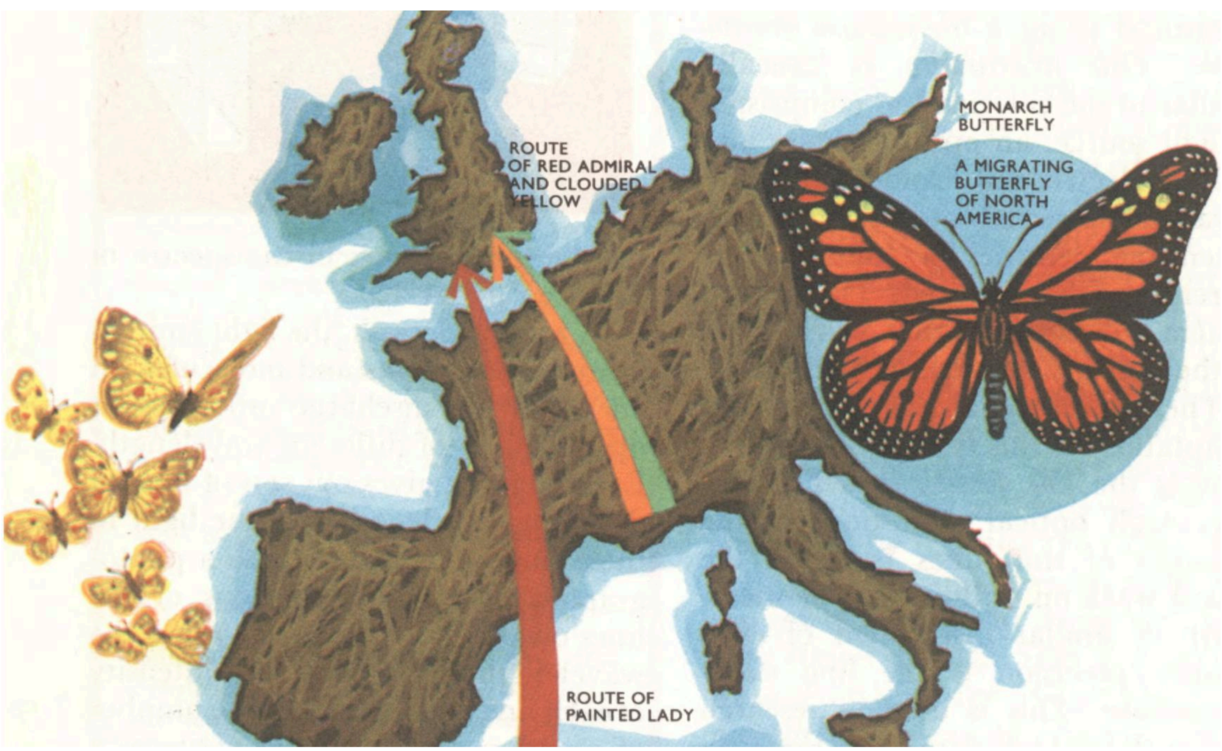


the information. Densities of Photographic Plates Another important application of optical density measurements is in the examination of photographic films and plates. When light falls on a photographic emulsion, the silver halide particles are converted into specks of metallic silver. These appear as the black photographic image when the film is developed. The optical density in the developed film depends on the exposure of the film —this is given by illumination x exposure time. It is of great importance to know the sensitivity of a film — how great an optical density is obtained when a fixed amount of light falls on it. In testing this property of the film, the manufacturer exposes the film to different amounts of light and measures the resulting densities. The film is exposed for a fixed length of time through a step-wedge filter—a series of optical filters of differing density that allow fixed amounts of light to pass to the film. The film is then developed and the optical densities of the different parts of the exposed wedge pattern are measured. In this way, a number of values of film optical density for different exposure values is found and a curve of density against exposure may be plotted. This curve provides a great deal of information about the film — not only its sensitivity but the contrast that can be obtained using the emulsion. The optical density of a film is LIGHT SOURCE PHOTOCELLS SENSITIVE TO DIFFERENT OLOURS LIGHT PASSES EQUIRED CELL INSERTED IN LIGHT BEAM THROUGH SELECTED L —— ae > ROTATION OF ; z a =, PRISM SELECTS p. Zs — } WAVELENGTH OF — ae | LIGHT USED MIRROR ve. MIRRO! measured using a transmission densitometer. This instrument is basically similar to the colorimeter, comprising a light source, an optical system and a photoelectric cell. Colour filters are invariably inserted into the light path when the densities of colour transparencies are measured. The optical » system is designed so that a small part of the film can be measured at a time. There are many variations »and adaptations of this type of instrument. One is the line densitometer. This has very high optical definition, so the densities of thin lines (such as the sound track on a film) may: be measured. A similar instrument of even greater precision is the line microdensitometer. This is used in research and analytical laboratories'to measure Transmission density measurements are made in testing photographic emulsions and in checking the optical densities of the different parts of exposed plates used in colour printing processes. The principle is very sumilar to that in colorimetric instruments —a photocell collects the light transmitted by the plate, in series with a selected filter. 4 i | PHOTOCELL — | — ay 32s : SAMPLE FILTER : the densities of lines in the spectra of different materials. In a spectrograph, the light emitted from excited atoms and molecules (in an arc or gas discharge tube) is split up into lines of different wavelength. Each element gives out sets of lines of different wavelength and the light in these lines is used to expose a photographic plate. The positions of the lines on the plate are used to find the wavelength of the light. The intensity of each line is a measure of the number of atoms present that are giving out light of that particular wavelength. If, for example, an alloy contains zinc and nickel, then the densities of the zinc and nickel lines on the plate can be used to determine the relative proportions of the two metals.


THIS DIAL ADJUSTED UNTIL ZERO READING IS OBTAINED ON METER
DENSITY READING
METER
PRISM POSITION IS CHANGED BY DIAL ROTATION
READINGS ON THIS DIAL INDICATES WAVELENGTH OF LIGHT

BIOLOGY Migration and Emigration AN'S commercial enterprise brings about the transport of goods from one part of the world to another. Unwittingly he has become a very influential agency in the dispersal of animals and plants. For instance, creatures from all sorts of countries may be brought to Britain in or on imported raw materials — timber, cotton, grain, wool. This sort of movement is completely accidental. Other such accidental dispersal mechanisms ROUTE OF PAINTED LADY sm, : it flies northwards into temperate zones. Here reproduction takes place. The following winter the bird returns to the tropics. The advantages of the migration are easy to see. Winter cold, in the north, with accompanying shortage of food is avoided. So are the torrid summers of the tropics. Even more vital are the migrations of amphibians. Frogs, toads and newts live most of their lives on land, but to breed, they must return to water. £ Doone Butterflies migrate from North Africa as well as birds. Routes are shown of the Painted Lady, Red Admiral and Clouded Yellow. The reason for the migration is not clear. Nor is the method by which the butterflies navigate. Inset: Monarch butterfly, a well-known migrating butterfly of North America. are high winds and floating logs. Far more definite, even predictable, shifts of animal populations are brought about by certain behaviour patterns of the animals themselves. There is migration—a fairly regular to-and-fro movement from one area to another and back again; and there is emigration, in contrast, a sudden random movement of animals — the animals do not return to their original territory. Migration The swallow winters in tropical and sub-tropical climates. In summer 2148 Consequently, each spring, a migration takes place from surrounding areas to local pools and rivers. The distance of the migration, in contrast to that covered by the swallow, is often very small. Migrations are especially common amongst creatures living in polar regions, or on high mountainous slopes. The harsh winters are avoided by movement into more sheltered places. The caribou, spending the summer wandering in the tundra, with the onset of winter move into the taiga — the northern coniferous forests. Migratory and emigratory behaviour — is instinctive. No conscious thought is involved. The action is entirely automatic. In the higher animals the pattern of behaviour is induced by changes in the hormonal system, particularly by influence of the pituitary gland. Some outside stimulation or stimulations are normally required to affect the hormonal system.

Increasing or decreasing length of day may be important, together with associated climatic variations. Vertical migrations in planktonic animals may be directly caused by differences in light intensity, temperature and salinity. The stimulus for emigration seems to be the overcrowding of a species with the consequent shortage of food and space, and the appearance of social stresses and abnormalities of behaviour. Mountain sheep, goats and chamois vacate the rocky upper slopes and move to lower levels. The greatest traveller of all is the arctic tern. This bird flies seasonally from arctic to antarctic regions — a distance of twenty odd thousand miles. Less commonly known migrants are the hermit-crabs which each year travel to the sea to breed and then return to land. Fishes are also great migrants; cod seasonally move to the north-west Atlantic to spawn; salmon and sturgeon pass from the sea into freshwater rivers, while the European eel may migrate 3,000 miles from Europe to the Sargasso sea to spawn. Migrations are not _ necessarily seasonal. Some planktonic animals Euphasia of the Antarctic retains its geographical position by migration each 24 hours. During the night these animals drift northwards in the current of cold water. Then they sink 600 feet and are moved by a current of warm water replacing the cold ANIMALS SINK

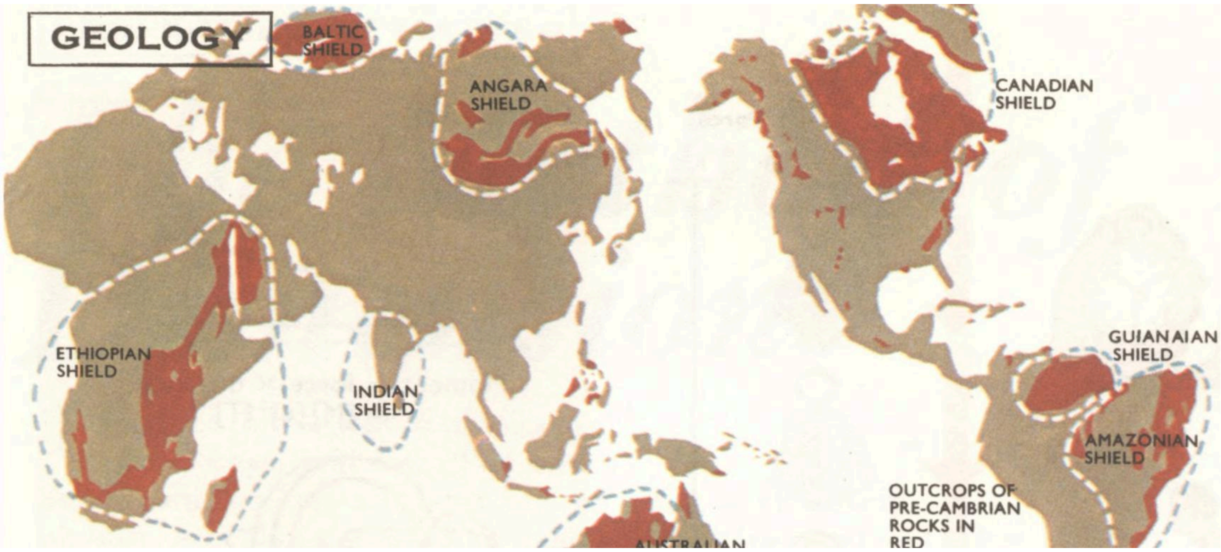


drifting in surface waters have a diurnal rhythm —a migration which takes place once every 24 hours. The organisms do not move horizontally. Instead they move from one level of water to another. The reason is rather obscure, for the migration takes place in different species at different times. Some creatures, for instance, sink during the day, rising at night; others the reverse. A possible long-term advantage is the continued change of feeding ground. One hundred feet down, the water is slow-moving compared with the faster-moving surface waters. On returning to the surface the organism has effectively changed its surroundings—for fresh surface water has replaced the surface water dators. Coming to rivers, seas or lakes, they move heedlessly on usually to self-destruction. But some emigrations in the past have met with spectacular success. The brown rat, now common in Europe, was unknown there before 1773. Then a huge invasion took place from the east. Descriptions of the swarms of rats swimming the great European rivers survive. Millions were drowned but some lived to continue the march. Soon the whole of Europe was occupied. The brown rats' successful entry into Europe can be closely compared with the 'suicidal' runs of lemmings today. The rats' invasion probably met with little competition from other creatures. Alternatively, the invasion coincided with slight BEES SWARMING The emigration of the brown rat at the end of the 18th century widely dispersed the species all over Europe. Though millions drowned in swimming the great rivers separating Asia from Europe, many survived, to breed. Right: a swarm of bees represents emigration in the insect world. Overcrowding of a hive leads to a division of the colony. in which the organism was originally swimming. Emigration In contrast, emigration by animals are population movements without return. Also called irruptions, they are intricately related to the size of the population of a species. A build-up in numbers producing overcrowding seems responsible for the outward drive. The lemmings of the arctic are famous for their periodic emigrations. Arctic voles and snow-shoe rabbits behave similarly, and away from polar regions so do some antelopes (such as the springbok) and some birds such as the waxwing. Usually emigrating creatures are doomed. Their headlong flight is accompanied by hysteria — an almost complete loss of self-preservative instincts. Apparently without fear, they are easy prey for following preclimatic changes making the new territories more acceptable to the brown rat. Thus emigrations are not by nature doomed from the start. Under certain conditions they mean, not mass destruction, but the colonization of new territories; the species becomes dispersed over wider areas. Even when emigrations do end in disaster, the species is favoured as a whole, for those few that do not migrate will become better fed and more healthy and more likely to breed successfully. Emigrations are not confined to the more advanced vertebrate animals. Similar movements are found in bees (swarming), locusts, dragonflies, ants, butterflies and termites. Nomadism Nomadism is a_ permanently wandering mode of living found alS. AMERICAN ALLIGATOR MIGRATES TO MAIN RIVERS WHEN TRIBUTARIES DRY UP 0 FROGS RETURN TO WATER FOR BREEDING ALS ONGREGATE FOR BREEDING SOME BA MIGRATE IN SEARCH OF FOODS All sorts of animals migrate. Migration may lead to an increase in food and ensures a species is spread over, and is exploring as large an area as possible. Migration alternatively may bring about congregation for breeding purposes, of animals which are by habit scattered over a wide area. most exclusively amongst the higher animals. Lower animals nearly all have a strong ‘homing instinct’ and though they do move around, they invariably return to a permanent place of ‘residence’. Nomadism is found particularly in mammals of desert regions or sparse plains. The vegetation being insufficient to provide enough food for a permanent population, the animals are forced to constantly change their feeding grounds. Nomadism is also found amongst amphibians and reptiles. Amongst birds it is rare; birds can travel such immense distances that though ranging over vast territories, nevertheless they can easily reach their homes. The vertical migration of the majority of plankton seems to ensure a change in surroundings. During each 24 hours, animals sink. When they return, currents have completely ‘changed’ the water with its food supplies... ae

PROPERTIES OF MATTER The Method of Dimensions HE three dimensions in physics are mass, length and time. The basic units, common to all branches of physics, are also the units of mass, length and time. All other kinds of units are derived units: analysis shows that, however involved the unit, whether it is a unit of force, viscous drag, heat or even electric charge, it can still be broken down into an expression (or a function) involving mass, length and time. The unit of area is a square unit, or The three dimensions, common to all branches of Physics, are mass, length and time. a length multiplied by a length. Area has the dimensions of [length]?. Volumes are measured in cubic units — a length multiplied by a length multiplied by a length. The dimensions of volume are therefore [length]*. (Square brackets are always used to denote dimensions.) The dimensions of mass, length and time are symbolized by [M], [L], and [T]. The unit of density is the gram per cubic centimetre, so it has the dimensions of [mass] divided by [length]® (per always stands for 'divided by'). 2150 ity) = ML [density] = IL Or, alternatively [density] = [M] x [L]* Whenever index numbers, for example the square index number 2 or the cube index number 3, appear 'under the line' as the divisor, it is exactly equivalent to minus the same index number on top of the line. The unit of acceleration is the centimetre per second per second, or [L] [T]* but it can also be written as [L] [T]~?. The multiplication signs are usually omitted. When the dimension is just [L] or [M] or [T], it is the same as [L}*, [MJ]! or [TJ]? The dimensions of force are particularly important because forces have to be considered in all branches of physics. The dimensions can be derived from [mass] and [acceleration]. Force is equal to mass multiplied by acceleration, so [force] = [M] x [L] x [T]°°. The dimensions of force lead on to the dimensions of energy, another unifying quantity, since it appears in every branch of physics. Energy in dynamics is a force multiplied by a distance. The dimensions of energy must be identical, wherever it appears, so [energy] = [force] x [distance] [energy] = [M] [L] [T]* x [L] This expression can be simplified, because the two dimensions of length can be collected together [L] x [L] = [L]* x [L]*= [L]*** = [L]', which demonstrates one of the useful properties of the indices. To multiply — simply add the indices, to divide — subtract. In mathematics, equations have just numbers (or symbols, like x and y, which represent unknown numbers) on either side. Unknown quantities appear in equations, in physics, but there is the additional proviso that [acceleration] = Area is [length]? and volume is [length]?. the indices of dimensions must be identical on both sides of the equation. It is no use equating a force on one side of an equation to energy on the other side —the dimensions are different. If the dimensions are different on either side the answer will inevitably be wrong. One of the equations of motion is written as s = ut + jat? s is a distance, u is a velocity, a is acceleration and t is a time. A quick check shows that the dimensions are correct. The left hand side of the equation has the dimension of length, so each of the parts on the right hand side must also have the dimensions of length. [ut] — [velocity] x [time] = [L}] [T]™ x [T]' [T]* and [T]~? cancel each other out, leaving only [L]. at? — [acceleration] x [time]? = (L] [TI]? x (T}* = [L}. So the whole equation is dimensionally in order. But perhaps the most endearing quality of dimensions is that they can be used to 'derive' quite involved formulae. No knowledge of the dynamics of oscillations is required to work out how the period of a simple pendulum depends on its length, and on the pull of gravity. The method of dimensions gives the relationship in a simple and elegant way. What can the period depend on?

Velocity ot wa Acceleration Velocity is distance per unit time — [leng unit time per unit time — [length] x [time]~ [mass] x [length] x [time]. Commonsense suggests that it may depend on the mass and length of the pendulum, and also on ‘g’, the acceleration due to the Earth’s gravity. But does it depend on (Mass)? or the square root of Mass ((Mass)!) ? Could it be (length)?, or simply (length)?? And how does ‘g’ enter the relationship ? The index of each of these quantities is unknown. The index [mass] is given the symbol A, the index of [length], B, and the index of [g], C. In other words [time] = [mass] x [length]® x [g]° [T}) = [M}* x [L}* x [[L][T 7° (we know that g has the dimension of acceleration [L] [T] 4. There are three ‘unknowns’, A, B and C, and there are three dimensions, [M], [L] and [T]. [M], [L] and [T] can be equated individually, giving three quite separate equations for the indices of [M], [L] and [T]. Equation for [M] L.hus. r.h.s. ar eee So the mass of the pendulum does not figure in the relationship. Equation for [L] Lh.s. r.h.s. o = B+C Equation for ['T] l.h.s. r.h.s. i = —2 x C | x [time]~}. Acceleration is distance per . Force is equal to mass times acceleration — From this equation, C = —4, and it follows from the previous equation that B = +4. So the final relationship is [T] = [M]° x [L]}' x [g} or [time] = [length]! x [g]~?. The index 4 means a square root, and the index —4 means dividing by a square root, so Deriving a formula — the Simple Pendulum | = i Mass | a. 4 . . = NO DEPENDANCE Mass is on one side only. LR 0 ON MASS Three dimensional equations give A, B, and C, but not the constant number which appears in the formula. This is an important limitation of the Method of Dimensions. __ [length]} gi [time] = 1 The a itself is not, however, equal [time] 1 rs The method of dimensions cannot give the pure number, 27, in the relationship. Only a full derivation from first principles can give this. to = it is in fact equal to 2m ji Length | x [s'|° THEREFORE B=-C=4

Present-day occurrences of PreCambrian rocks throughout the world. FIRST SIGNS of LIFE MAGICAL line in the geologic time scale separates the Cambrian period from earlier, PreCambrian times. The first Cambrian deposits— they were formed about 600 million years ago-—are rich in fossils, the preserved relics of previous life. Representatives of 14 of the 16 major groups making the animal kingdom are known to have been flourishing; so were most groups of plants. But in the Pre-Cambrian rocks, despite their great thicknesses (for Pre-Cambrian times occupy at least four-fifths of the whole geologic timescale) fossils, as a rule, are exceedingly rare. What are these fossils and why are they so scarce? What light do they throw on the subsequent evolution of life? Until 1947, intensive searching of Pre-Cambrian rocks in all parts of the world had brought forward meagre evidence of life in these remote times. PRESENT-DAY SOFT CORALS THE 'SEA-PENS' Present-day 'seapens' or soft corals. Forms like this have been found in PreCambrian rocks in England, South Africa and Australia. I 'CANADIAN 'SHIELD GUIANAIAN =e SHIELD » oe Probably the oldest signs of life are graphitic (carbonaceous) deposits. These are found scattered in many localities throughout the world and some of them are at least 2,500 million years old. Certainly, in more recent geologic times, carbonaceous deposits are known to have been derived from plant and animal remains. Probably of equally remote age, are certain algal-formed limestones found in South Africa. Estimated at about 2,000 million years, are certain microscopic fossils found preserved in black flint nodules in Ontario, Canada. The fossils include two types of algae, two types of fungi and a flagellate (i.e. simple protozoan animal). In Montana, U.S.A., the Beltian limestones contain large fossil reefs, some over 20 feet high and 30 feet wide. The reefs are believed to have been formed more than 1,000 million years ago, by certain types of algae. None of the actual algal structures remain however. The Montana limestones have also yielded fossils of sponge spicules and what may be fossil brachiopods. From the Pre-Cambrian rocks in the Grand Canyon vicinity of the U.S.A. come possible sponge spicules and perhaps the impression of a jellyfish. In Finland, minute tubes of carbon may be the remains of PreCambrian plants; radiolarian remains come from Brittany in France and in Charnwood Forest, England, and South Africa, 'sea-pens' or soft corals have been discovered. Possible worm tracks impressed in rocks have been found in Pre-Cambrian rocks of all sorts of ages, and the abundant deposits of Pre-Cambrian ironstones may have their origins in the reducing activity of certain bacteria. In 1947, a late Pre-Cambrian mud flat was discovered in rocks of the Ediacara hills, South Australia. Pressed into the ancient deposits were the shapes and forms of innumerable soft-bodied creatures. Some are remarkably similar to modern forms. There are many-segmented annelid worms resembling today's ocean-going forms, numerous jelly-fish and more 'soft'? corals — the sea-pens. There are undoubted worm trails preserved and several types of animal unknown both today and elsewhere in the fossil record. The Appearance of Skeletons The treasure-house of PreCambrian fossils in Australia, together with the meagre evidence from elsewhere, has done much to illuminate what really happened at start of the Cambrian times. The old idea that — ees Geolo Bis. cateatiapaleey


GEOLOGY
BALTIC SHIELD
ANGARA SHIELD
CANADIAN SHIELD
ETHIOPIAN SHIELD
INDIAN SHIELD
GUIANAIAN SHIELD
AMAZONIAN SHIELD
OUTCROPS OF PRE-CAMBRIAN ROCKS IN RED

The first Cambrian rocks, with their comparative wealth of fossils, were deposited about 600 million years ago. All older rocks were formed in Pre-Cambrian times. As the world is believed to have been formed 4,500 million years ago, theoretically, Pre-Cambrian times lasted 3,900 million years. But actually the oldest known rocks existing at the Earth's surface are 2,700 million years old, leaving 1,800 million years of the Earth's history unaccounted for.

Pre-Cambrian rocks occur in every continent. Often they are deeply buried but in many areas, uplift and erosion has exposed them at the surface. In fact, in Canada and in Australia they occupy most of the surface. As a rule Pre-Cambrian rocks are hard. If they were not initially igneous in origin (cooled from a molten mass), they were hardened and altered by heat and great pressures (metamorphosed rock).

So far, two very broad divisions of

Pre-Cambrian rocks are recognised. The *Archaeozoic* (dawn of life) consists of rocks much older than 2,000 million years and includes deposits with possible traces of life. The *Proterozoic* (first life) incorporates all later Pre-Cambrian rocks.

Pre-Cambrian Climates Despite the remoteness of PreCambrian times, the climates then do not seem so unlike our own today. Thick accumulations of limestones perhaps represent deposition in warm seas while mudcracks preserved in ancient sediments indicate seasonal drought. Periodically, in some places, torrents of rain fell sufficient to wash masses of coarse sediment into nearby freshwater fakes. Perhaps most surprising of all, the Pre-Cambrian period had at least one Ice Age of its own and perhaps more. Boulders scratched b moving ice and glacial deposits are weal preserved, particularly in Australia and India. the groups of animals found in Cambrian times did not exist any earlier is obviously wrong. Certainly they were flourishing in late Pre-Cambrian times, if not before. Another idea — that the animals did exist in much the same form but have just not been preserved — is not convincing either. Quite a variety of Pre-Cambrian rocks are known which could have preserved hard parts of organisms. Far more likely is the theory that Pre-Cambrian organisms were all softbodied. Very exceptional conditions, as in the Ediacara Hills, would be needed for any chance of preservation. If, too, many of the organisms swam or floated in deep water, there would be even less chance of preservation, for deep-water sediments into which their remains would fall, are rare in geologic record. The enormous increase in number and types of organisms in Cambrian times would then seem to coincide with the relatively rapid evolution of hard parts in many stocks. As materia] for building skeletons, e.g. calcium The Pre-Cambrian period occupies fourfifths of all geological time. The rarity of fossils makes it difficult to further sub-divide this interval. Below, some soft-bodied worms and jellyfish found in late Pre-Cambrian rocks of Australia. Evolution of hard parts in a relatively short space of time seems to have brought about the relative abundance of Cambrian fossils. tA aAd-4! rp, UNIDENTIFEGS mS ‘7 A ORGANISM ts oe “SS > ae — a e £ “fe, a — - , . SS oath JSULYFISH = an oe ore e "Sed PRE-CAMBRIAN © FOSSILS FROM™ EDIACARA HILLS, ae AUSTRALIA carbonate, seems to have been abundant in Pre-Cambrian seas, the emergence of shells etc. most likely represents an advance in biochemical evolution. The time needed is not as short as the ‘sudden’ appearance of hard parts in the fossil record would indicate. ‘Sudden’, in geologic terms, usually represents several millions of years. The forming of hard skeletons giving support and protection would enable animals to adopt a variety of new Traces of early Pre-Cambrian life include algal secretions, worm trails, sponge spicules and the possible mould of a brachiopod. LIMESTONE iff NODULE A SECRETED “BY ALGAE ee, «WORM

| MATHEMATICS | Mapping into Logarithms HERE is no limit to the variety of sets of numbers. The only requirement of a set is that each of the numbers in it (the elements of the set) has some property in common with all the other numbers. The basic set in mathematics is the set of natural numbers le Ae | When these are spaced evenly along a number line, addition becomes equivalent to moving right along the number line, and subtraction is equivalent to moving left. Other important sets are based on the powers of one of the natural numbers. The first power of a number is the number itself. The second power is the number multiplied by itself. For the third power, the number is multiplied by itself another time, and so on. | The Set of Powers of 2 Power, or log. to base 2 Index notation Number 2154 The powers of 2 are I2 2X2 2X2xX2 2X2X2X2... I 2 44 8 16 A short-hand notation is used for writing higher powers. 2 x 2x 2x22 (32) is written as 2° (two-to-thefifth). 5 is called an index number, and denotes how many factors of 2 make up the number. Another name for it is the logarithm to the base 2. Logarithms are a very useful mapping of the natural numbers. Numbers can be represented by their logarithms in much the same way as a building can be represented by a black block on a map. The set of buildings is mapped to the set of blocks {buildings} —» {black blocks} and logarithms a tnumbers}" = {8 numbers = — ae | The Set of Powers of 10 Power, or log. to base 10 Index Number notation 10° 10' 10? 10° 10* Very large numbers are often written using powers of ten as a ‘shorthand’. For example, the number of molecules in one gramme-molecular weight of a substance (Avogadros’ number) is written 6-02 x 107°, rather than 602,000,000,000,000,000,000,000. => ONE-TO-ONE CORRESPONDENCE CORRESPONDENCE e 7} sy @ MANY-TO-ONE oe One number in the number set corresponds to a single number in the logarithm set — there is satd to be one-to-one correspondence. The number can be mapped to the logarithm set for multiplication and dwwision, and then mapped back to its original set. This would be impossible if a logarithm corresponded to several numbers (many-to-one correspondence), because we would not know which number to choose when mapping back to the number set. or number set: 2 4 8 16 32 64 yyy v4 log. set: bee 84: 58 Each of the numbers in the logarithm set corresponds to a single number in the number set. This is called one-toone correspondence. If the numbers are spaced out evenly along the number line, their corresponding logarithms

will not be evenlyspaced. If the logarithms are evenlyspaced, their corresponding numbers are compressed into small, unequal spacings on the number line. But, in this second case, the logarithms can be added simply by moving along the number line. Adding a logarithm is not the same as adding its corresponding number. In the process of mapping from a number set to a logarithm set, the mathematical operations have changed. Multiplication in the number

Mapping the Arithmetical Operations

$$4 \times 16 \leftrightarrow 2^2 \times 2^4 = 2^6 = 64$$
$$\log_2 4 \times \log_2 16 \leftrightarrow 2 + 4 = 6 = \log_2 64$$

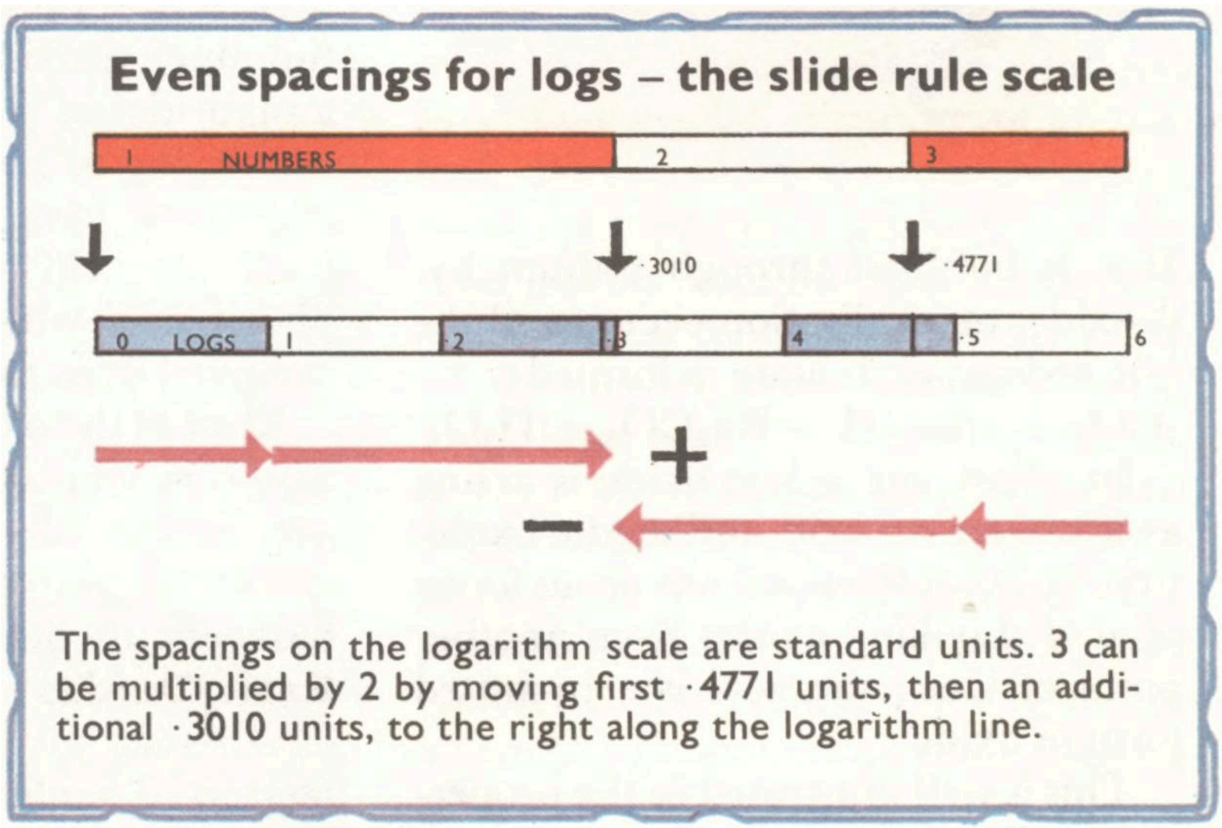
therefore $\times \rightarrow +$

$$64 \div 32 \leftrightarrow 2^6 \div 2^5 = 2^1 = 2$$
$$\log_2 64 - \log_2 32 \leftrightarrow 6 - 5 = 1 = \log_2 2$$

therefore $\div \rightarrow -$

Adding the indices (the logarithms) is the same as multiplying the numbers. Subtracting the indices is equivalent to division. But the method works only when the numbers to be multiplied or divided belong to the set of powers of the same number.

set has been mapped into addition in the logarithm set, and division has been mapped into subtraction. Addition and subtraction are usually much easier operations than multiplication and division. So complicated multiplication and division sums can be mapped into the set of logarithms to make the calculation easier. Then they are mapped back from the logarithm set to give the answer in terms of ordinary numbers. Why does multiplication map into addition? 4 x 8 in the number set is 2% x 2°. 4 x 8 is 32, or 2°, which is also 2?**. Multiplication is achieved by-adding the indices, or logarithms. 8 divided by 4 is 2. These numbers in the index notation are 2°, 2? and 2. The answer, 2! is the same as 2? — in other words, division has become equivalent to subtraction in the logarithm set. But it is not quite so simple as it may at first seem..Fer-instaficepat is not possible.to" multiply 2* and 3* by ae arranged along a number. pai +, —y HK Fle rule is aN n eMibers on the 'x' scale are not spaced evenly, but positioned so that their corresponding logarithms (to base 10) are evenly-spaced. So the slide rule can be used for multiplication and division by adding and subtracting lengths of wa ial \Even spacings for numbers — no use on a slide rule F010 : eee i Jt Numbers which are evenly-spaced can be added and subtracted by moving to right and to left along the number line. Here the corresponding logarithms (to base 10) take up segments of increasing size. There is no standard unit size. this method. 3% is not a power of 2 and 2* is not a power of 3. The numbers to be multiplied do not belong to the set of powers of the same number. Numbers can be successfully mapped into logarithms for multiplication and division only if they belong to the set of powers of the same number. Only a very small proportion of numbers are powers of two, so using logarithms-to-the-base-two has only limited application. It is, of course, possible to base logarithm sets on the powers of three, the powers of four, the powers of five and so on. But always the usefulness of the mapping is restricted to the set of powers of each of these numbers, and the result of multiplying or dividing always belongs to this same set. To simplify the situation, practical logarithms are based on ten. Although £ only 1, 10, 100, 1,000, and-so on can i be written as whole-number powers of ten, all numbers can be written as fractions (or decimals) of powers of numbers, each 6 THIS SCALE CANNOT BE USED FOR ADDITION IT Is TOO UNEVEN logypA + logy 9B log o(*3010 *4771 . 3 is 10 "4771ten. 2 is 10 2 eae 10°! x 10 or {numbers} 2 x 3 , ¥ 4 (numbers) “3010 + -477 The sign —> means ‘corresponds to’. The result of the addition is -7781. This is logy96. 6 is 10°778?. However, in devising the systems of logarithms to the base 10 (logy), the original definition of a logarithm as a power of a number has changed. It is better to decide on a new definition of logarithms. This is based on their actual properties. If there are two numbers, symbolized by A and B, then log, 9A and log,9B are defined so that A x B). e 4 OE BY Yo ducliniitstit as 12 9100 129 6 ‘ulindiy Treas oo statu’ Se ee ~ OO Age ee $s oO The spacings on the logarithm scale are standard units. 3 can be multiplied by 2 by moving first -4771 units, then an additional -3010 units, to the right along the logarithm line.

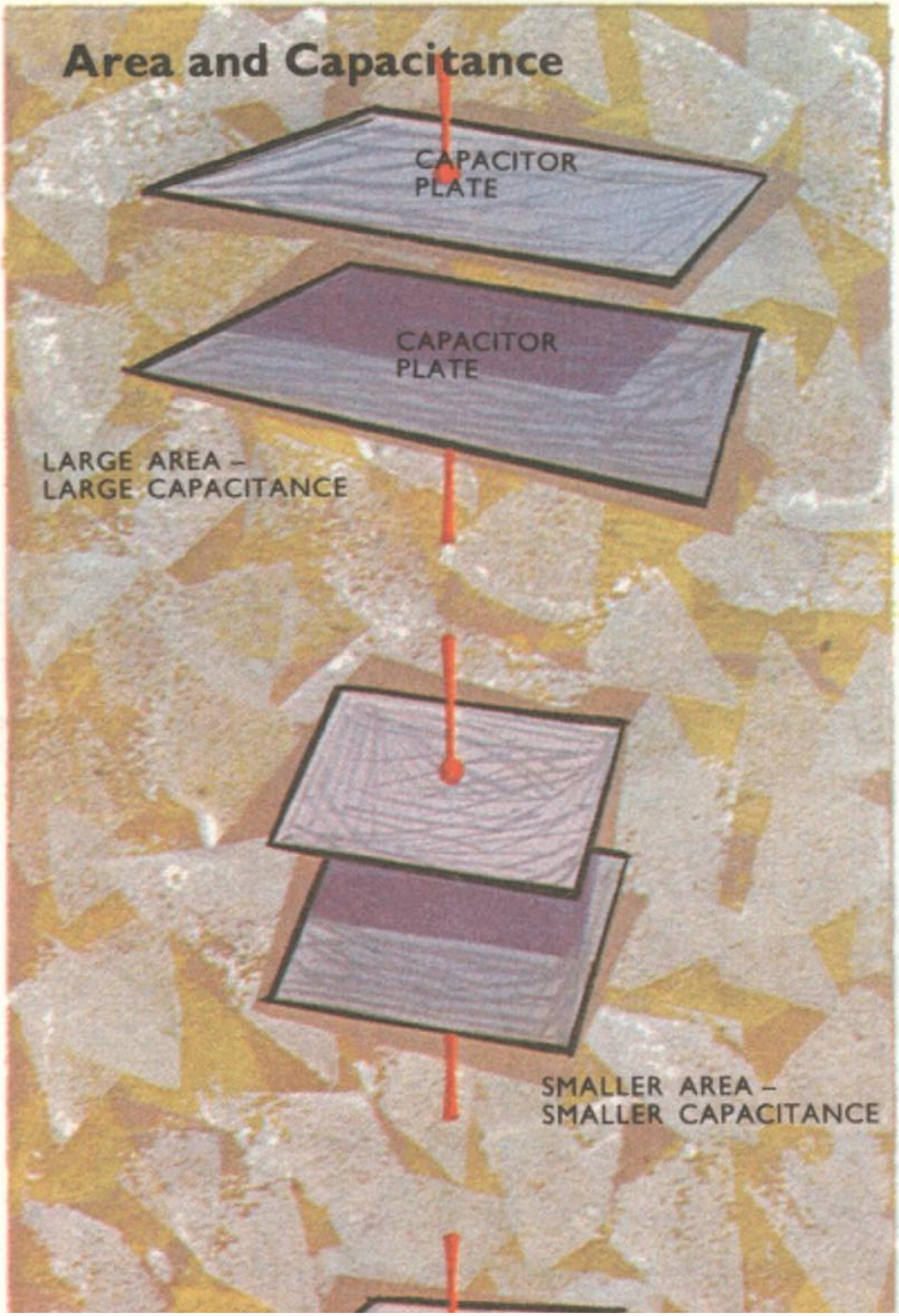


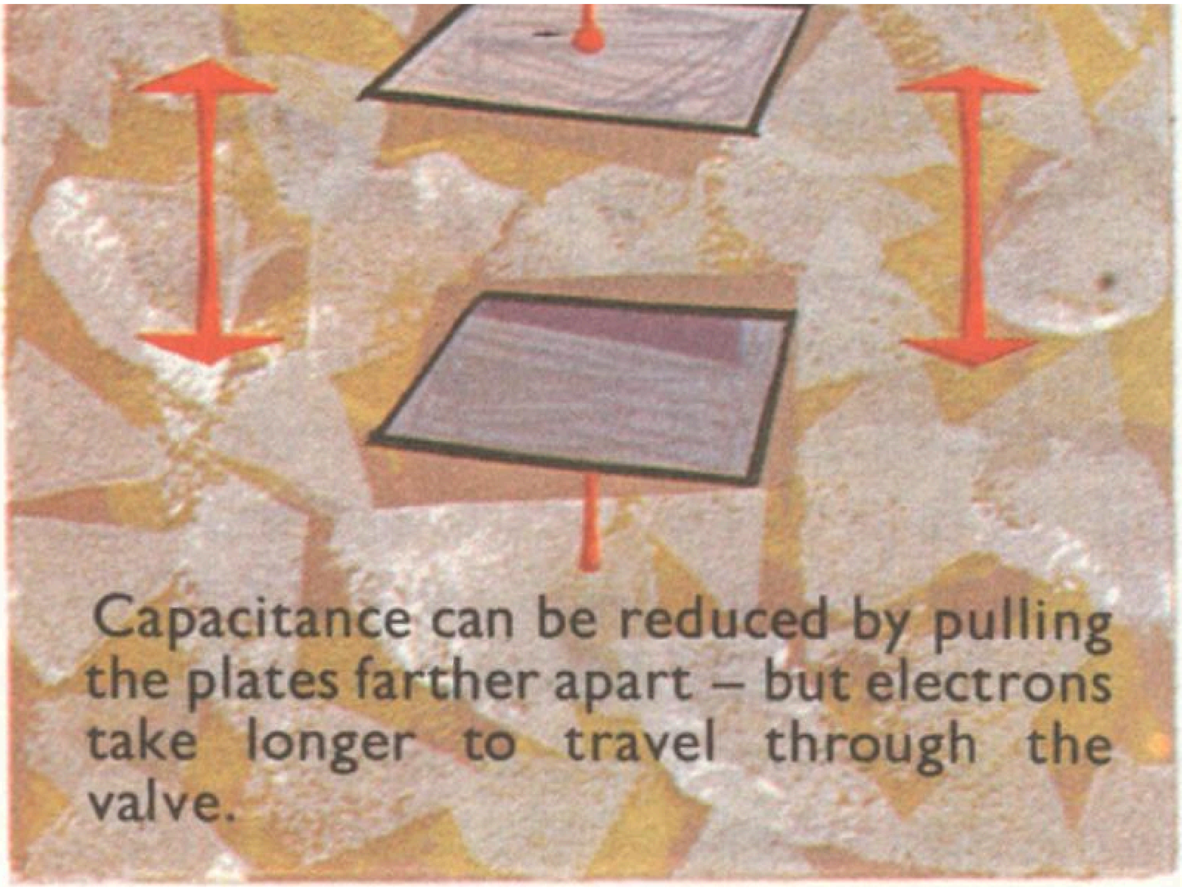
INORGANIC CHEMISTRY Sulphur dioxide ts an acid oxide. In water, it forms sulphurous acid. WHEN chemical elements combine with oxygen, they form oxides. These are a very important class of chemicals because they form the starting points for two very important types of reagent — acids and bases. In solution, acidic oxides become acids, basic oxides become alkalis. If carbon, in for example, coal or coke, is completely burned in air, it combines with oxygen to form carbon dioxide. Carbon dioxide is an acidic oxide. Like sulphur dioxide, it causes atmospheric corrosion because of its acidic properties. If carbon dioxide is dissolved in water, a weak acid, carbonic acid (HgCQOs) is produced. H,O + CO, —— H,.COs. If it is bubbled through sodium hydroxide, neutralisation occurs and the salt, sodium carbonate, is formed: CO, + 2NaOH > Na,CO 3 a6 H,O. In effect, the acidic oxide is acting as if it were an acid and acidic oxides may be considered as anhydrous forms of acid. Looking at this from another point of view, the acid is a hydrated form of oxide. This is well illustrated in the proper2156 oxide. It sometimes acts as an acid oxide _ (left), sometimes as a basic oxide (right). ACIDIC and BASIC OXIDES ties of the two of the more important oxides of sulphur, sulphur dioxide (SO,) and sulphur trioxide (SOs). When dissolved in water, sulphur dioxide forms sulphurous acid (H,SO3), but on heating, the sulphur dioxide is easily driven off. The formation of acid from oxide and oxide from acid is a reversible process, written as: H,O + SO, = H,SO; Sulphur trioxide, however, reacts violently with and dissolves in water Amphoteric Oxide Aluminium oxide is an amphoteri to form sulphuric acid: H,O a SO; ——- H,SQ,. This is not a reversible process and the water has to be forcibly extracted, using phosphorous pentoxide, to recover the oxide. 2H.SO, ats P,Oio > 2SOz3 ate 4HPO; Sulphur trioxide is, then, a true anhydride of sulphuric acid. It is interesting to note that the dehydrating agent, phosphorous pentoxide, is itself an acid oxide forming phosphoric acid with the water molecules removed from the sulphuric acid. Most of the oxides of non-metals are acidic in character — families of acids are produced when, for example, oxides of nitrogen, silicon and the halogens are dissolved in water. Basic Oxides Whereas the oxides of non-metals tend to be acidic in character, most of Sodium oxide is a basic oxide. In water, it forms the alkali sodium hydroxide. the metals form basic oxides. When they combine with water, alkalis are formed. For example, sodium oxide dissolves in water to form sodium hydroxide: Na,O + H,O +> 2NaOH. Basic oxides are also formed by calcium, copper and iron, and all form hydroxides with water. Amphoteric Oxides Most oxides are either definitely acidic or basic. But there is an interesting group of oxides that behave as acids in the presence of alkalis, and bases in the presence of bases. These are called amphoteric oxides. They are ormed by a number of metals — aluinium, tin, zinc, and antimony amongst them. Aluminium oxide, for example, forms aluminium chloride in the presence of hydrochloric acid, and thus acts as a base: Al,O; + 6HCI + 2AICl; + 3H,O. In the presence of sodium hydroxide, however, it acts as an acid, forming sodium aluminate: Al,Os3 = 2NaOH > 2NaAlO,; a H,O This reaction is used in the extraction of aluminium from its ore, bauxtte.

BIOLOGY The MECHANIS of theSTOMA (GREEN plants make their own food. They build up simple chemical compounds such as carbon dioxide and water into complex food (i.e. starch). Since carbon dioxide is usually taken from the air, gas exchange is needed between the atmosphere and the plant's internal airspace system. If the outer impervious layer of the plant — the epidermis — was continuous over the whole plant surface, such a gas transfer would be impossible. But scattered amongst the thick-walled, tightly-packed epidermal cells are small elliptical openings surrounded by two bean-shaped cells — the guard cells. A pair of guard cells and the opening between them makes up a structure called the stoma (plural, stomata); it is through the stomata that carbon dioxide diffuses into the plant. In letting in carbon dioxide, the plant also loses water vapour — a substance vital to its existence. When Turgor in opposing guard cells forces them away from one another. The stomatal pore in between, becomes enlarged. Collapse of the cell causes the adjacent walls to come together again so eliminating the pore. PALISADE TISSUE AIR SPACE water is plentiful the loss is not important for uptake replaces the water which has evaporated. But in drought, water loss needs to be drastically reduced even if carbon dioxide supply has to be cut down at the same time. The stoma is an automatic device built into the leaf with just this function. The mechanism is quite simple even though all the agencies by which it works are not fully understood. The guard cells on either side of the opening are modified epidermal cells. Their distinctive crescent or bean-shape has already been mentioned. At each end they are firmly joined together but are separated centrally leaving the slit (or stomatal pore). Differing from ordinary epidermal cells, the guard cells have unevenly thickened cell-walls — thin on the side away from the pore, becoming much thicker along the surface in contact with the pore. Chloroplasts, structures rich in chlorophyll —the organic catalyst needed for photosynthesis — are also present, 'o CHLOROPLASTS GUARD CELL OPEN enabling guard cells to build up food. When water is plentiful, the guard cells are turgid — swollen up with the water. The thinly lined cell wall, on the side away from the opening, stretches. The thickly lined cell wall against the opening does not. The result? —the stretching of the thin wall pulls the two thickened walls apart away from the opening. The size of the pore consequently increases and the maximum amount of diffusion can take place. In dry conditions, water is lost from the guard cells. Turgidity is lost and, in becoming limp, the inside edges of the two guard cells come together, closing the opening. The quantity of water is not the only factor which influences the mechanism. Light is also important. Somehow insoluble starch synthesized in the cells is converted into soluble sugars. This increase in concentration means that more water flows in from neighbouring cells by osmosis. With darkness the reactions are reversed and the opening is closed by the collapsing guard cells. As a rule, stomata open in the daytime and close at night. Stomata on stems are sparse, for stems are not usually important for photosynthesis. More commonly they are concentrated on the leaves— usually on the under surface where they are not so exposed to drying agencies. Vertically growing leaves, for instance, of grasses, commonly have equal numbers on each surface, while plants with leaves that float, such as lilies, have their stomata completely confined to the upper surface. The actual number of stomata varies with plants, 65 per sq. centimetre on the upper surface of a daffodil - to 450 on the lower surface of an oak leaf. Pores, when completely open, rarely occupy more than 3% of the total leaf surface. But what is important, is that a number of minute pores let in more carbon dioxide than a single large one. The numerous minute - stomata allow almost as much carbon dioxide to diffuse into the plant as through a whole leaf surface. The » possession of stomata, however, does enable an effective water-saving mechanism to function — especially important in very dry weather. 2157

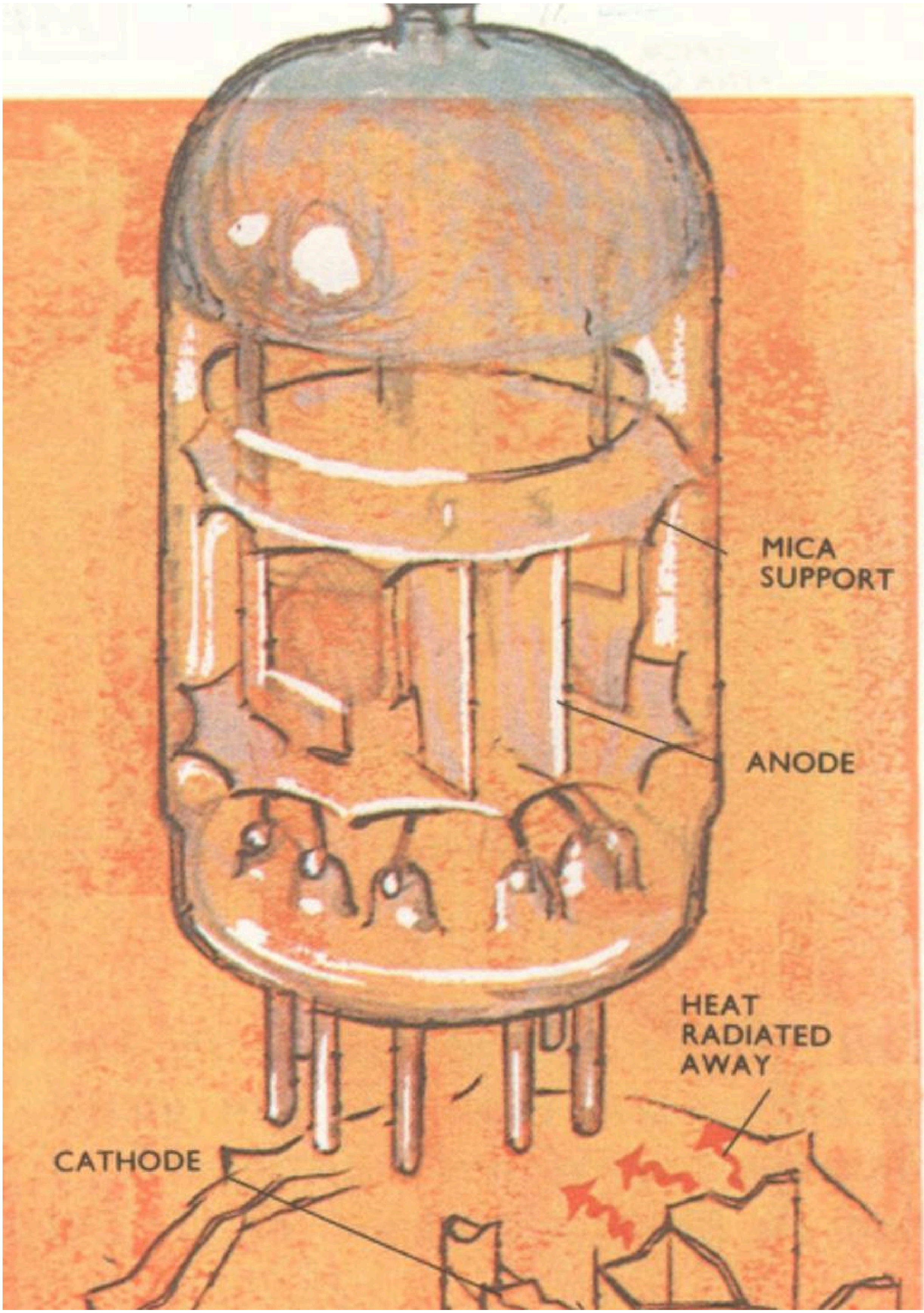
ELECTRONICS VALVES at HIGH FREQUENCIES— The BEAM TRIODE "TELEVISION transmissions start at about 50 megacycles per second. In other words, the lowest frequency of radio wave on which the television picture can be carried is 50 million cycles per second. The electronic valves in the receiving television set must be able to deal with these very high frequency signals. Unfortunately, 50 megacycles (mega always means million) is just about the frequency at which the ordinary triode valve (the electronic valve with three electrodes) stops working efficiently. It still amplifies the signal, but not as well. One of the main reasons for this is that two of the electrodes in the triode, the cathode and the grid, must be very close to each other. They are so close that they start influencing each other at high frequencies, acting like a very small capacitor. The higher its frequency, the easier it is for an electrical signal to pass across a capacitor. The capacitance starts to become a nuisance in a variety of ways. One of The pentode valve amplifies high frequency signals, but it is too 'noisy' to be used in the early stages of a television set. CATHODE CONTROL GRID SCREEN GRID SUPPRESSOR GRID ANODE | BELOW CROSS-SECTIONAL DIAGRAM OF A PENTODE | VALVE these is that part of the signal, instead of appearing as the output of the valve, and going on to the next stage of amplification, finds its way back, through the grid and cathode, to the input of the valve. The amplification of the valve drops and the valve stops being efficient. The problem becomes more and more serious when the frequency of the signals is increased. 50 megacycles per second represents the lowest of the television frequencies. As more and more television 'channels' are used, available lower frequencies are filled up, and higher frequencies are used, so the conventional triode valve has to be replaced. One way of tackling the whole problem of capacitance is to add extra grids — spirals of fine wire, between the original grid (the control grid) and the anode of the valve. An extra grid turns the triode into a tetrode, and a second extra grid, to counteract some of the unpleasant effects of the first extra grid, turns the valve into a pentode. The pentode is more efficient at high frequencies than the triode. But it is much more 'noisy'. The signal passes through the valve as a cloud of electrons from the cathode, and passes through the grids on its way to the anode. In a pentode the current has to find its way through two extra grids. Although the grids are fine spirals of wire, some of the electrons inevitably hit them. The electrons are jostled about. As a result, the electron current fluctuates slightly. The fluctuations produce unwanted ‘noise’. In a sound receiver this is an unpleasant rushing noise in the background. In a television receiver, the valve ‘noise’ is seen as a graininess in the picture. The valves in the first stage of a elevision set should be as noise-free as possible, because any noise proCATHODE GRID SUPPORT ODS GRID ANODE ONE ELECTRON BEAM HERE BEAM HERE Pas’ This beam triode has three electrodes. Electrons from the cathode are directed in two beams towards the anode. The capacitance between cathode and grid and cathode and anode is largely eliminated. The first stage of a television receiver deals with the weak signal picked up by the aercal. Noise from the valves at this stage 1s amplified in subsequent stages. U.H.F. AERIAL ONE ELECTRON ELECTRONS HIT GRIDS AND GENERATE NOISE 2158 duced in the first stage is going to be amplified in subsequent stages. After the first stages, the noisiness of a valve is not such an important consideration. The triode is the least noisy valve. So the best way of overcoming the capacitance nuisance is to modify the triode itself, rather than to turn it into a pentode or tetrode. Most of the capacitance comes from the part of the grid with the most area, the rods which support the fine spiral wire grid. In the beam triode, the electrons in the valve are restricted to SOUND SIGNAL VISION (VIDEO) SIGNAL a beam which misses the rods completely. Their capacitance is eliminated, because they do not influence the current. In this way, the total capacitance between the cathode and the grid is reduced about three times. Increasing the spacing The closer the grid is to the cathode, the greater the effect cathode and grid have on each other -— that is, their | capacitance is greater. So a possible solution to the problem would be to space the grid farther away from the cathode. But, like many solutions, it creates a new problem. The electrons take longer to travel through the valve. The voltage on the grid is changing rapidly in time with the high frequency signal reaching the television set. As it changes, it varies the amount of current which can reach the anode. But if the electrons take as long to travel from the cathode to the grid as the signal voltage takes to change from its maximum to its minimum, they find when they get to the grid that conditions have completely changed. In the confusion, the amplification drops. Valves at higher frequencies A development of the beam triode is a triode with the electron beam in one part instead of two. In this way the area of anode facing the cathode is reduced — and so is the capacitance between cathode and anode. The anode has to conduct all the current flowing through the valve without overheating — it should have a large surface area so that the heat can be dissipated. A large surface area leads to a high capacitance, so the anode is shaped with a long thin ‘nose’ opposite the cathode. Only the thin part of the ‘nose’ contributes to the capacitance, while the rest of the anode helps to radiate the heat away. ANODE > GRID SUPPORTS An _ultra-high-frequency valve. The anode is ioniegneed) to reduce capacitance. The leads to the grid of the valve are earthed, and spaced to eliminate inductance. Noise produced in the first stage of the television set is amplified in following stages. It appears as a background rushing noise from the loudspeaker, and a graininess in the picture. JOISE’ 1S A SHUSHING BACKGROUND LOUDSPEAKER

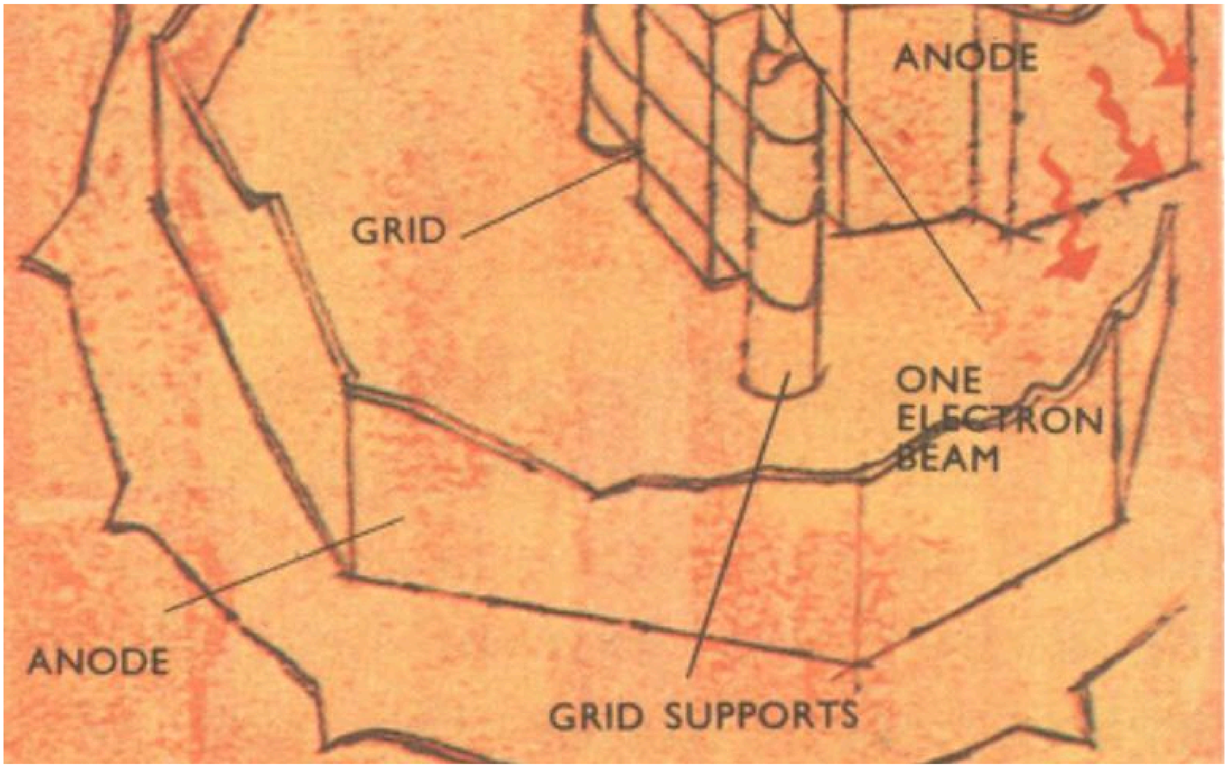
DEFLECTION COILS ‘NOISE’ IS A GRAININESS ON THE SCREEN - LIKE AN OVER-ENILARGED PHOTOGRAPH


Area and Capacitance
CAPACITOR
PLATE
CAPACITOR
PLATE
LARGE AREA –
LARGE CAPACITANCE
SMALLER AREA –
SMALLER CAPACITANCE

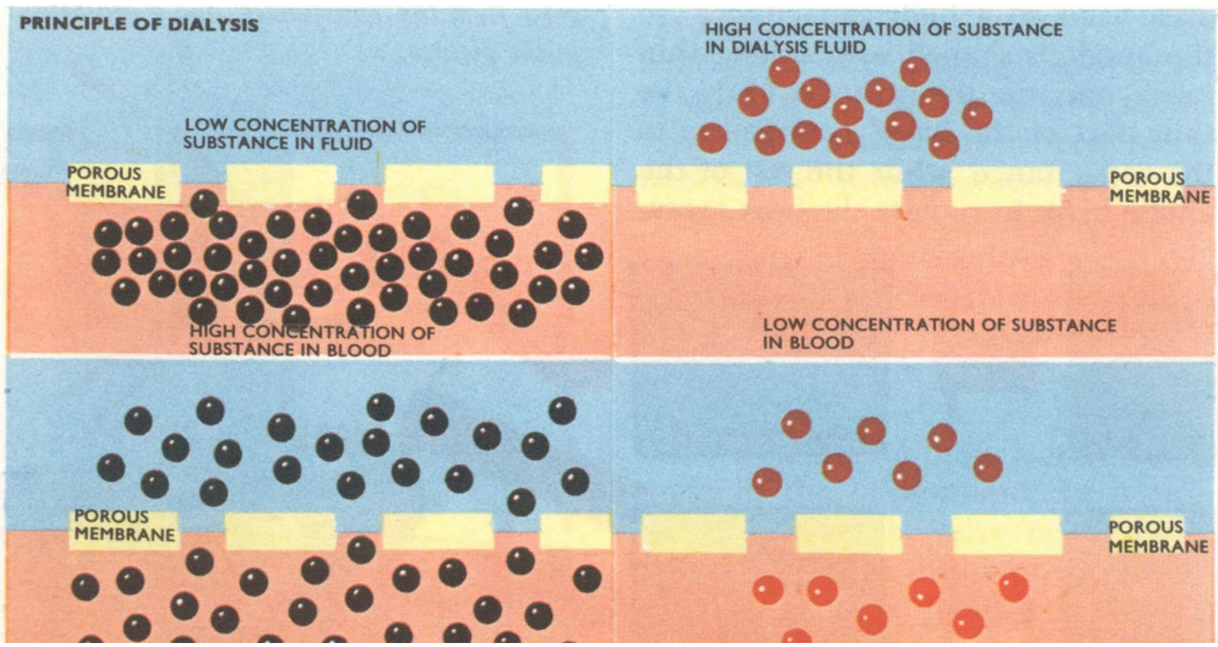


Capacitance can be reduced by pulling the plates farther apart – but electrons take longer to travel through the valve.


MICA SUPPORT
ANODE
HEAT RADIATED AWAY
CATHODE


ANODE
GRID
ONE
ELECTRON
BEAM
ANODE
GRID SUPPORTS

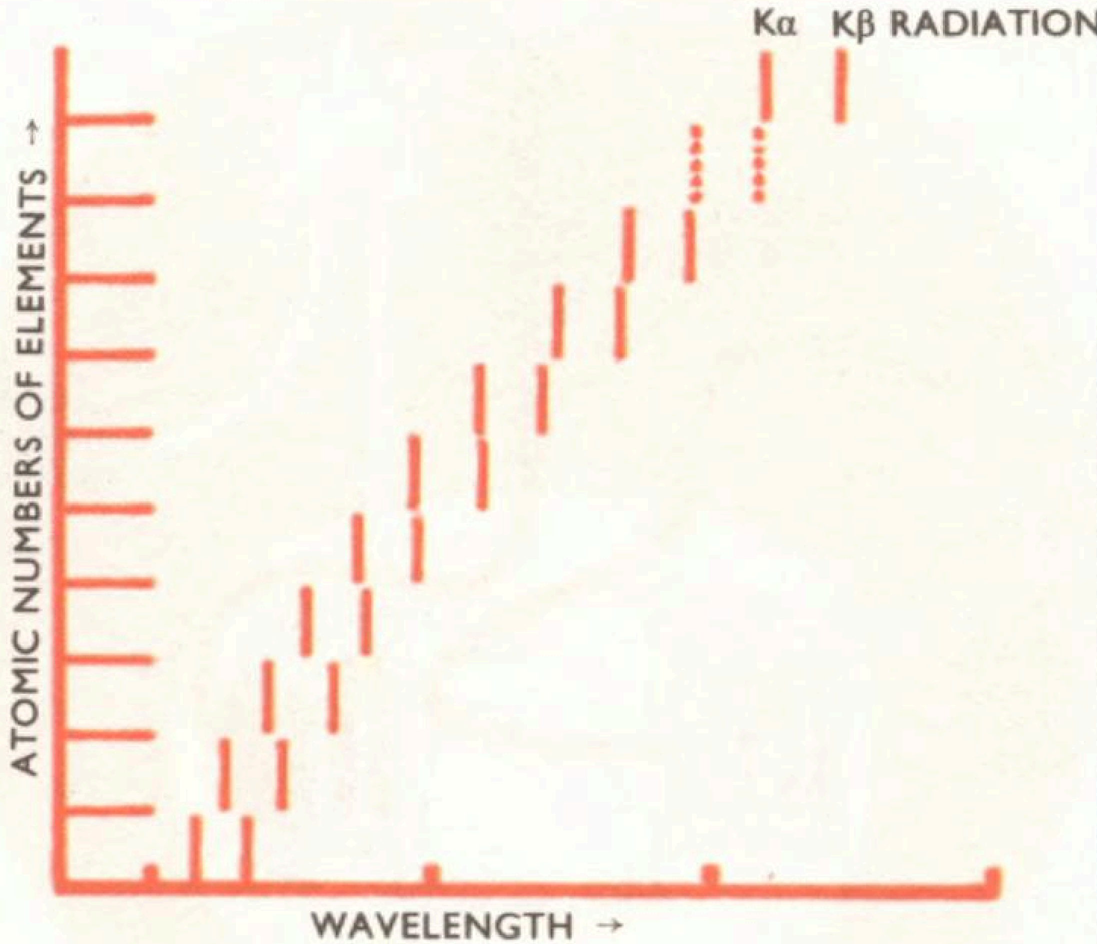
MEDICINE GLOMERWEUS BOWMAN'S CAPSULE Sat __ COLLECTING TUBE ARRANGEMENT OF KIDNEYS KID KIDNEY TUBULE Left: Man's excretory system: the pair of kidneys are fed with blood by the aorta and the blood is returned, purified, to the heart along the inferior vena cava. The two ureters carry away wastes from the kidney to the bladder. Middle: A single kidney tubule showing the Bowman's capsule and collecting tube. The blood vessels entwining the collecting tubule selectively reabsorb many of the substances filtered Srom the Bowman's capsule. Right: Detail of Bowman's capsule with glomerulus. The ARTIFICIAL KIDNEY T the back of the abdomen nestling under the diaphragm and embedded in protective layers of fat lie the kidneys—a pair of bean-shaped organs, each about the size of a clenched fist. The kidneys are as vital to Man as his heart or brain. They remove all sorts of wastes from the blood, which would poison the body if allowed to accumulate, and they keep the complex composition of the blood within narrowly defined limits despite the variety of physiological activities recurring throughout the body. Though a kidney looks solid enough, really it is made up of about a million minute hollow structures. Each consists of a cup-shaped head (a Bowman's capsule) and a long winding tubule leading from it. All the tubules eventually join up to form the ureter, which leads, in turn, to the bladder. Every Bowman's capsule is provided with blood; minute capillary vessels form a knot — the glomerulus — actually within the capsule. Blood pumped into the glomerulus is at a far greater pressure than fluid contained outside, Showing the principle of dialysis — how substances move through pores in a membrane from an area of high to an area of low concentration. By preparing a suitable dialysis fluid components can if necessary be removed from the blood (right). ABOUT EQUAL 2160 POROU: MEMBRANE LOW CONCENTRATION OF SUBSTANCE IN BLOOD CONCENTRATIONS ARE ABOUT EQUAL in the rest of the capsule. As the walls of the capillaries are porous, much of the blood's components, including nearly all the water, are forced through into this fluid; all that is left are blood cells and proteins which are too large to pass through the pores. But as the filtered-out constituents move down the tubule, nearly all the water and 'many of the dissolved substances are reabsorbed into the blood stream: not all, however. In some manner not fully understood, the quantities of different substances are selectively reabsorbed by the blood stream. Foodstuffs such as glucose are entirely returned but various electrolytes in the blood such as sodium and bicarbonate ions are taken up in quantities which vary with the requirements of the body. The poisonous waste products such as urea, on the other hand, are not _ reabsorbed, but pass to the bladder,dissolved in a minimum of water and eliminated from the body. The Artificial Kidney Obviously the artificial kidney cannot hope to match the natural kidney's delicate functioning, its perfect coordination with other parts of the body. Nevertheless, introduced in 1945, it has become an important life-saver and an instrument of limitless research. Its essential component is just a long tube made of a cellophane membrane. ELY REABSOR| OOD



Cellophane resembles the walls of blood capillaries in having pores of almost exactly the same size. Consequently, the same types of substance can penetrate both surfaces. The principle on which the artificial kidney functions — dialysis — is also very simple. If two solutions are separated by a porous membrane, molecules and atoms will move to and fro from one solution to another. For any one type of particle, the overall transfer will be towards the solution with least concentration of that particle; the particles will tend to move from one area of high concentration to an area of low concentration. To remove poisonous waste products such as urea, blood simply has to be circulated in cellophane tubes through a solution in which the concentration of this substance is low. As the movement of each type of atom or molecule is largely independent of the other types, a 'rinsing' fluid can be prepared of right composition not only for removing waste products such as urea but keeping the concentration of other substances at a constant level, and if necessary, actually adding ingredients. The earliest type of kidney functions solely by dialysis. The instigator of this machine was Wilhelm Kolff. Blood from a patient with defective kidneys is passed along a cellophane tube wrapped about a rotating drum. The rotation of the drum moves the This mini-coil artificial kidney is capable of ultra-filtration. The design is based upon Kolf?s disposable coil kidney, first developed in 1956. SUPPORT DIALYSIS FLUID OUTFLOW BLOOD IN blood by gravity, through a bath filled with rinsing fluid and noxious substances are removed without upsetting the balance of other substances. Such artificial kidneys have saved hundreds of lives from uremia (selfpoisoning due to a build up of urea in the body) when natural kidneys through shock or disease have failed to function properly. Of far wider application today are artificial kidneys designed to not only remove wastes by dialysis but also to remove excess body fluid from the blood by ultrafiltration. A very common variety is based upon Kol ff's disposable col kidney developed in 1956. Here, unlike the rotating drum models, coils of cellophane tubing are supported by fibre-glass screens and wrapped about a central cylinder. The whole unit fits neatly into a container through which rinsing fluid is in continuous circulation. The compression of the cellophane tubes by the screens maintains INFLOW through. Under this pressurt, excess fluid in the blood flows out through the membrane. A pumping device can be used to boost the pressure of the blood before it passes through the coil. Other Uses for the Artificial Kidney he pressure of the blood flowing Accumulated waste products such as urea can be removed from the body using the artificial kidney. The abnormally high or low concentration of any other ion or compound can also be rectified. All that is needed is a rinsing fluid containing the correct concentration of substances. Thus the artificial kidney provides a_ standard ENVELOPE OF POLYVINYL CHLORIDE DISTENDED BY CIRCULATION OF DIALYSIS :* LUID DIALYSIS FLUID INFLOW ROTATING COUPLING j DIALYSIS > f FLUID dl ; iry> ROTATING / COUPLING / REVOLVING DRUM Principle of the rotating-drum artificial kidney. Celluloid tubing 1s wrapped about the drum ; when the drum rotates the blood is moved by gravity through the bath of prepared dialysis fluid. Exchange of tons purifies the blood. method for treating barbiturate or salycic poisoning. In this respect the artificial kidney is more efficient than the natural kidney, for the poisons can be completely removed from the blood stream; there is no question of any reabsorption after filtering. Since the artificial kidney can effectively control the quantities of the various constituents of the blood, it is a simple matter in experiments to change the concentrations of specific substances whilst at the same time keeping other constituents at a constant level. In this way the causes of a number of diseases can be discovered with means of applying remedies. Keeping the Blood Flowing When the artificial kidney was . developed one immediate difficulty to overcome was how to prevent blood clotting once it was circulating in the cellophane tubes. Originally hirudin was used. Hirudin is a substance produced by leeches — the Airudinea. These creatures lead a blood-sucking mode of life and the hirudin facilitates their feeding. In order to extract enough of the chemical for an experiment, thousands of leeches were necessary. Today other anticoagulants, which can be prepared artificially, are available. Most important is heparin, discovered in the early 1930's. 2161

INCIDENT ELECTRON ATOMIC PHYSICS 8 @ a) cs) e e ee Ses @ @"»6 @ oe: "ELECTRON REMOVED 8 @ ELECTRONS ARE ® . 3 ARRANGED IN K, L, M, N SHELLS > INCIDENT 3 D ELECTRON BOUNCES OFF M RADIATION N RADIATION X-Rays Spectra HE use of X-rays is now quite common-place. Within a few weeks of Rontgen's historic discovery, in 1895, that these invisible rays could penetrate matter and expose photographic film, they were being used to help surgeons locate and diagnose fractures. Since that time, their practical uses have grown enormously, not only in medicine, but in industry. But to the fundamental scientist, the study of X-rays has provided valuable information about the nature of matter itself. Not only has it given him insight into the way that atoms and molecules are arranged in crystals —it has also answered many questions about the structures of atoms themselves. How X-rays are Made X-rays are a form of radiant energy, very similar to light and radio waves, but of much shorter wavelength. They are thrown out by an atom when the energy of an electron 'missile' is absorbed. In an X-ray tube, electrons are emitted by a heated, negativelycharged cathode and are accelerated to bombard a_ positively-charged anode. The speeds attained by the electrons depend on the voltage difference between the two electrodes. If a voltage difference of 100,000 volts is applied, electrons can be accelerated to half the speed of light. In the atom, there is a tiny, positively-charged nucleus, surrounded by a cloud of electrons. These electrons are arranged in energy shells. Electrons in the inner shells possess the lowest energies — they need to be given much more energy to be removed from the atom than those in the outer 2162 shells. Whatever shell an electron is in, it needs a definite amount of energy to remove it from the atom. When a very high energy electron is absorbed by the atom an electron in an innermost shell is displaced from the shell. This state of affairs does not last very long, because the displaced electron is immediately replaced by a higher-energy electron that normally resides in the next shell out. Then, an electron from Ka KB RADIATION ATOMIC NUMBERS OF ELEMENTS > WAVELENGTH > In 1913 Moseley found a simple relationship between the positive charge on the nucleus of an atom (the atomic number) and the wavelengths of the 'K' X-rays given out by the atom. In its simplest form it showed that the wavelength of the K radiation increased in 'jumps' as shown in the diagram, going from one atom to the next in the Periodic

Table. Where there was a gap in the diagram there was an undiscovered element, and new elements have been discovered, and their atomic numbers found, by studying X-ray spectra. Even more important, Moseley's work established the importance of the atomic numbers of elements — that the essential difference between one element and another was the difference in numerical positive charge on the nucleus. Producing X-rays. The target 1s bombarded by fast-moving electrons and an inner (K) electron is removed. An electron from the next (L) shell takes its place, and its place in turn is taken by an M electron. Similar transitions can occur inall the shells soa mixture of K, L, M, N radtation, of different wavelengths, 1s emitted. the next shell moves in to take the place of the second electron, and its place is taken by an electron from the next shell out. Each of the electrons' 'journeys' from one shell to the next makes the atom radiate energy. of a definite wavelength. The transition to an inner, higher energy shell gives out (hard) radiation of high energy and short wavelength. Radiation caused by a replacement in an outer shell will produce longer wavelength (soft) X-rays. The X-ray Spectrum The radiations involving electrons in the different shells have been named according to the traditional names given to the shells. For example, radiation arising from the removal of an electron from the inner 'K' shell will result in radiation of a particular wavelength. This is called K radiation. Radiations from the next L,M,N, shells result in the longer wavelength L,M,N, radiations, so when a target is bombarded with electrons, a mixture of radiations is thrown out. But the shortest wavelength radiation that is emitted cannot have a greater energy than that of the electrons that bombard the target. Only high energy electrons can produce K radiation. Electrons of lower energy will produce a mixture of L,M,N, radiations. Just as the light emitted by an ordinary lamp can be split up into its different wavelengths by a prism in a spectrometer, so can the X-rays in an X-ray spectrometer. In the Bragg spectrometer, the beam of X-rays is reflected from a flat crystal. In the crystal are layers of atoms, one after the other, and X-rays are reflected from these different layers. (Strictly speaking, this is a diffraction effect, rather than reflection. The waves are not reflected as by a mirrer but are scattered in all directions by the atoms in the crystal planes.) The X-rays reflected from one layer 'mix' with X-rays from the

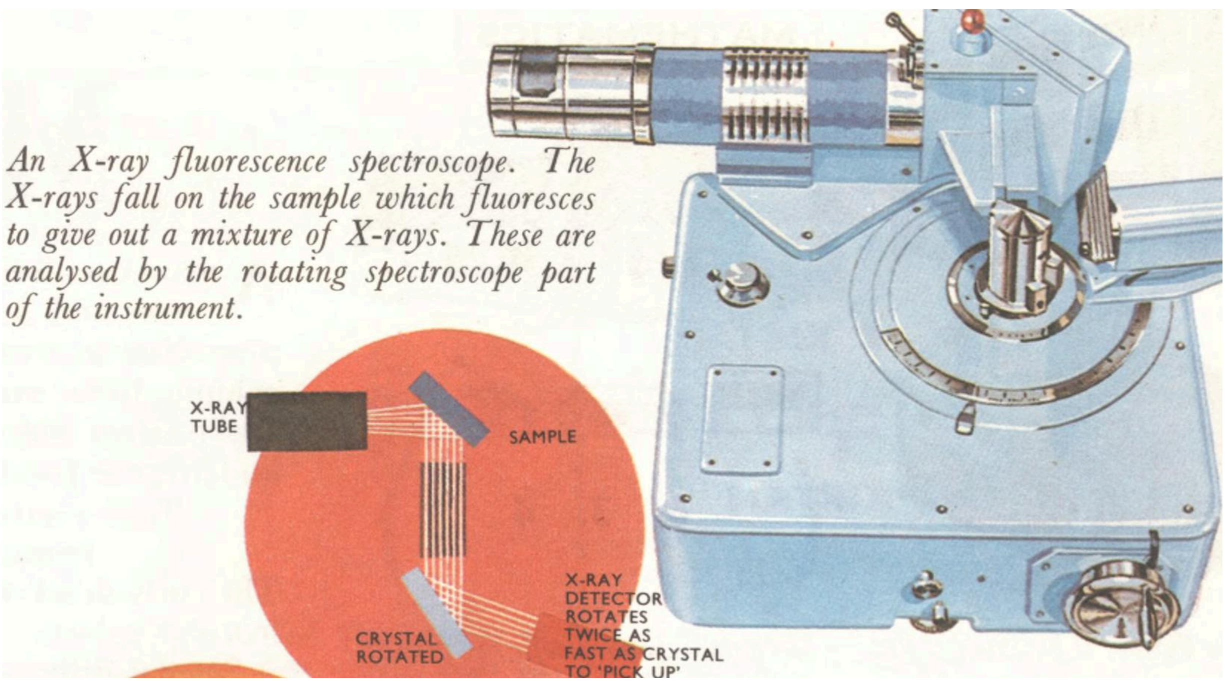


In 1913 Moseley found a simple relationship between the positive charge on the nucleus of an atom (the *atomic number*) and the wavelengths of the 'K' X-rays given out by the atom.

In its simplest form it showed that the wavelength of the K radiation increased in 'jumps' as shown in the diagram, going from one atom to the next in the Periodic Table. Where there was a gap in the diagram there was an undiscovered element, and new elements have been discovered, and their atomic numbers found, by studying X-ray spectra. Even more important, Moseley's work established the importance of the atomic numbers of elements – that the essential difference between one element and another was the difference in numerical positive charge on the nucleus.

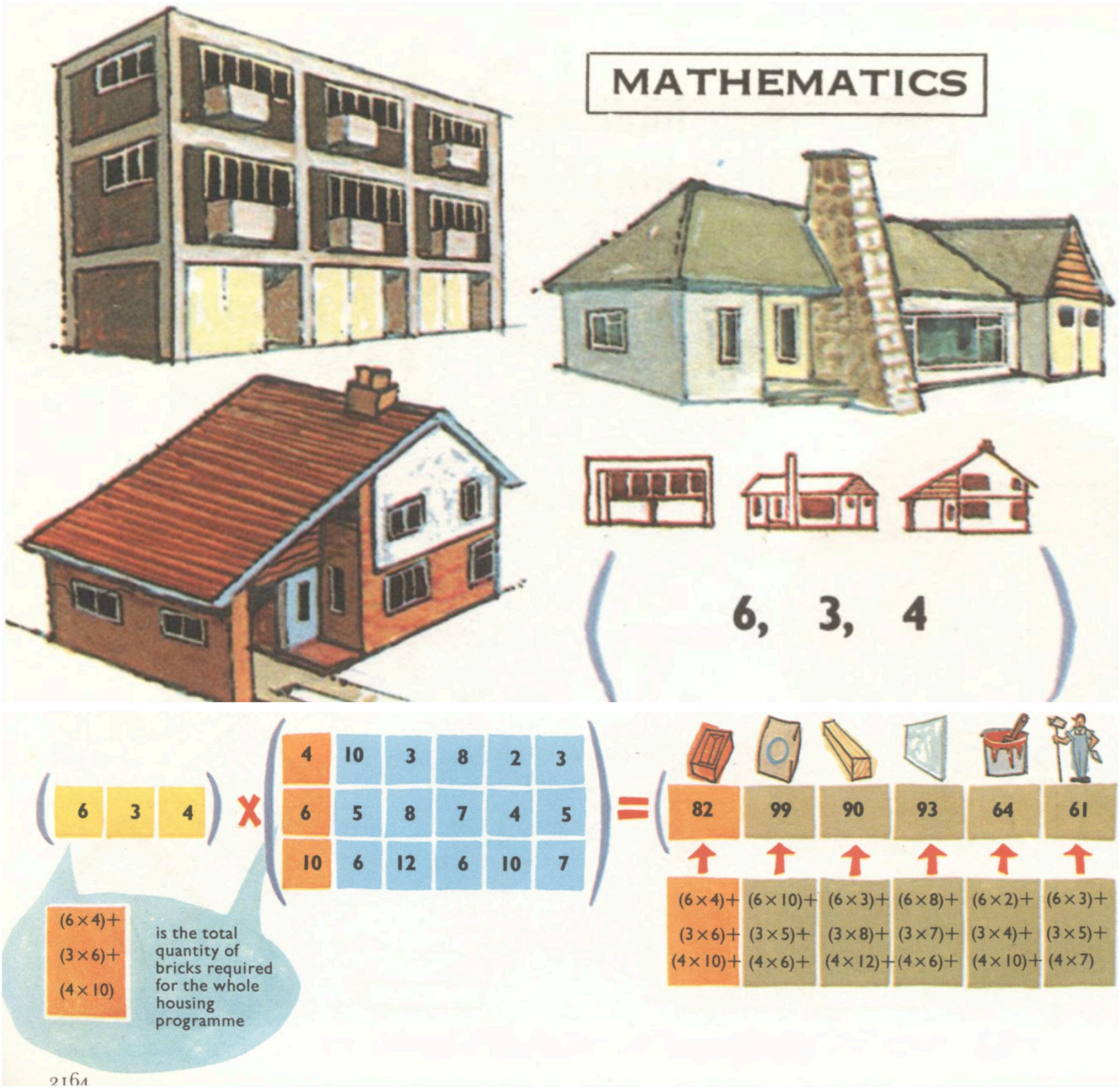
next. At certain angles of reflection the X-ray ‘waves’ travel exactly one complete extra wavelength in passing to the second layer, and back. Then, the two sets of reflected waves are added together, because ‘crest’ coincides with ‘crest’ for a particular angle of reflection, so with X-rays of a particular wavelength, there is a strong reflection. If the crystal is rotated in the X-ray beam, the different angles of reflection each represent ‘strong’ reflection angles for a different particular wavelength of X-rays. Suppose the target in an X-ray tube is giving out a mixture of radiations (K,L,M,N,). By rotating the crystal, a series of strong K,L,M,N, reflections will flash into view. The strengths of the reflected rays can either be measured using an ionization chamber or scintillation counter or can be registered on a photographic plate and then measured. By noting the variation in X-ray intensity with angle of rotation, the change in intensity with X-ray wavelength is found. In this X-ray spectrograph the various radiation ‘peaks’, corresponding to the K,L,M,N, radiations, are clearly seen. The different ‘lines’ (K,L,M,N etc.) are, in fact, not single peaks but are split into a number of single peaks. This is because an electron can have a number of different energies within a shell. These give rise to a mixture of radiations. For example, there might be Ka,KB,Ky, peaks all appearing on the spectrograph, each representing an energy subdivision. An X-ray fluorescence spectroscope. The X-rays fall on the sample which fluoresces to give out a mixture of X-rays. These are analysed by the rotating spectroscope part of the instrument. CRYSTAL FLUORESCES, GIVING OUT X-RAYS Applications of X-ray Spectroscopy The existence of the various radiation peaks has provided valuable information to support the modern theory of the atom. In addition, the spectrograms that are obtained are always the same for atoms of a particular element, so a valuable analytical method, of great use to the chemist, is made available. In practice, there is a particularly valuable variation of this method as an The X-rays are split into different wavelengths by reflecting them from a flat crystal. (Left.) The different X-ray spectral lines are formed on the photographic film, for different positions of the crystal. (Right.) The X-ray intensity-wavelength graph is ‘drawn’ by measuring the intensities of the reflection at different angles using the ionization chamber. X-RAY TUBE AT DIFFERENT ANGLES RYSTAL ROTATED IVES STRO! PHOTOGRAPHIC PLATE EXPOSED TO PROVIDE LINES AT

‘STRONG’ REFLECTION ANGLES NG REFLECTIONS Beene ED RADIATIONS analytical tool. This is in X-ray fluoroscopy. It is very similar to the phenomenon of optical fluorescence. Atoms can be made to give out visible light if they are stimulated by light of the appropriate wavelength. Atoms also give out X-rays if they are stimulated by X-rays of the appropriate wavelength. A very intense beam of X-rays falls on the unknown substance, from a high energy X-ray tube. The photons of different energy in the beam cause electrons in the atoms of the substance to be displaced and X-rays are emitted. The wavelengths of the X-rays are analysed using a rotating crystal spectrometer. The spectrum can be shown on a photographic film and the method provides a very quick and reliable method of analysis. The positions of the peaks on the photograph indicate what elements are present. The sizes of the peaks indicate the quantities of each element present. ROTATED WITH CRYSTAL (AT DOUBLE ANGULAR SPEED) 2163

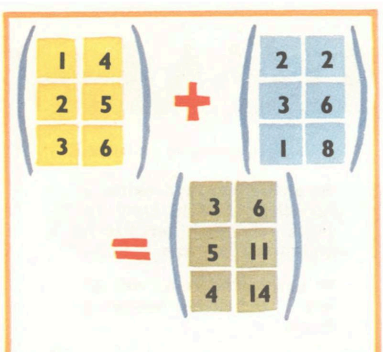


An X-ray fluorescence spectroscope. The X-rays fall on the sample which fluoresces to give out a mixture of X-rays. These are analysed by the rotating spectroscope part of the instrument.

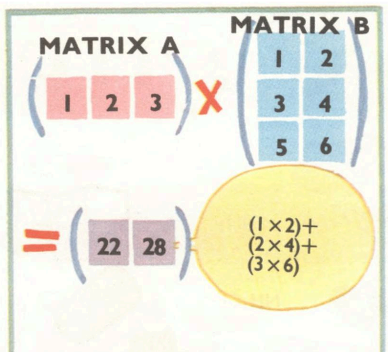
The builder is building 6 flats, 3 bungalows, and 4 detached houses. This can be represented by a single-rowed, three columned matrix, (6, 3, 4). BS Here, the basic house-building materials are simplified to six — bricks, cement, wood, glass, paint and labour. Each type of house requires a different amount of each material. The matrix showing the estimates of materials and labour for each house. Numbers are arbitrarily chosen to give an indication of relative amounts only. To find the total amount of each material which the builder must order for the whole housing programme, this matrix is multiplied by the (6, 3, 4) matrix. Matrices AQ matrix 1s a rectangular array of numbers, arranged in rows and columns. The numbers in each row have something in common, and so have the numbers in each column. In the example illustrated, row numbers represent the relative amounts of each basic material used for building one particular design of house. {Basic materials mort (Row A of the) (house A (matrix | The curly brackets mean 'the set of' and the arrow means 'correspond to'. Several different house designs are being costed. The amounts of each of the basic materials in each type have been estimated, and these vary from design to design. The numbers in the columns of the matrix represent the relative amounts of each basic material in each type of house. ee material "| . age X of es in each design matrix If the builder is making only one house of each design, the totals of each of the columns are the total amounts of each of the materials he will need for the whole building programme. But different numbers of each type are ordered. If 6 of design A are being built, 3 of design B and 4 of design C, numbers in the first row (row A) must be multiplied by 6, row B by 3 and row C by 4, to give the total amount of material necessary. Sorting out which numbers should be multiplied together is largely a matter of common sense. However, the matrix, besides organizing the initial information, helps to organize operations such as addition and multiplication. The 6 of design A, 3 of design B and 4 of design C form a three-columned, single-rowed matrix (6, 3, 4) The total amounts of materials are found by multiplying numbers from this matrix by numbers from the larger matrix. Obviously not all pairs of numbers are multiplied together. The rules of matrix multiplication set out which When a stngle-rowed, three-columned matrix is multiplied by a three-rowed, six-columned matrix, the result is yet another matrix — with a single row, and six columns. This is one of the basic rules of matrix multiplication. (1 row, 3 columns) x (3 rows, 6 columns) = (1 row, 6 columns)



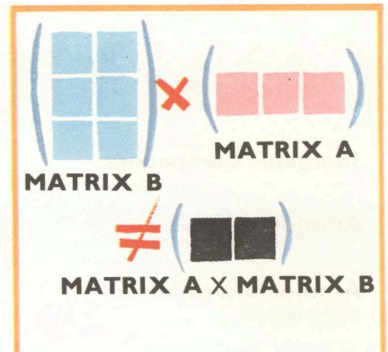
Before two (or more) matrices can be added (or subtracted) they must have an identical number of columns and rows. second. pairs are to be multiplied. The first column in the 'requirements' matrix is the number of bricks in six flats, plus the number of bricks in three bungalows, plus the number of bricks in four detached houses. Three pairs of numbers are multiplied, and then the results added together, to give the first column in the requirement matrix. Costing each house involves another matrix multiplication. The cost-per-unit of each of the basic materials is represented by a six-rowed (if there are 6 basic materials) single-columned matrix. The total cost of building each design is found by multiplying the initial matrix by this matrix. Multiplication is possible only when the number of columns in one matrix is equal to the number of rows in the other matrix. The restrictions on adding together two matrices are even more stringent. The number of rows and columns must be identical in each matrix, and then each of the Multiplying matrices. The number of columns in the first matrix must be equal to the number of rows in the @ MATRIX A MATRIX B +(am) MATRIX A X MATRIX B Multiplication does not work when the matrices are multiplied the other way around — matrix A X matrix B is not the same as matrix B atrix A numbers making up the final matrix is equal to the sum of corresponding numbers in the two initial matrices. For example, the builder might be building houses in two different regions, and putting different amounts of the basic materials to suit local requirements. So he has two different three-rowed, six-columned matrices, and the result of adding them together is another three-rowed, sixcolumned matrix giving the amount of material needed to build one house of each type in each region. Matrices are used when there are several different variables (the building materials) put to several different uses (the three house designs). Another typical example might be the amounts of the vitamins A, B, C, D and E in four different foods. This could be represented by a fourrowed, five-columned matrix. Total food requirements, and the cost of a balanced diet, can be found by multiplying this matrix by a requirement matrix, and a cost matrix. Now the builder works out the total cost of each type of house. The cost matrix has six rows, one column. When the original three-rowed six-columned matrix is multiplied by this, the result is a three-rowed, single-columned matrix. Each row in this matrix is the total cost, to the builder, of one of the three house designs. The figures quoted here are representative amounts, and bear no relation to actual costs. yer se The cost matrix for building materials RD ££ “eae enrllC erTTlllC el eC wn EEE ee eee ee eel \ j THREE ROWS ONE COLUMN — RELATIVE COSTS



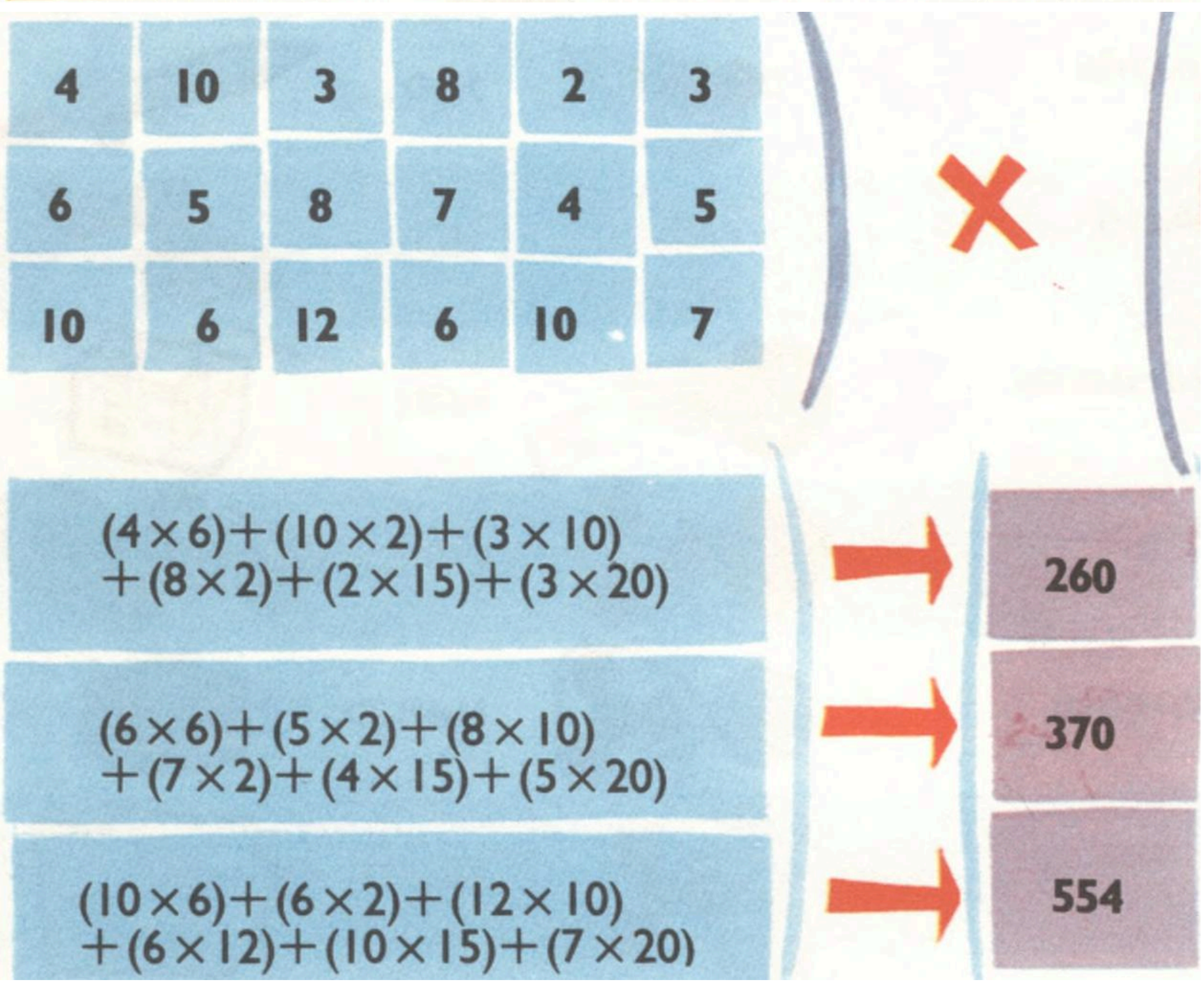
Before two (or more) matrices can be added (or subtracted) they must have an identical number of columns and rows.



Multiplying matrices. The number of columns in the first matrix must be equal to the number of rows in the second.



Multiplication does not work when the matrices are multiplied the other way around – matrix A × matrix B is not the same as matrix B × matrix A.



| APPLIED SCIENCE | COMMON CHEMICALS Inorganic Compounds Ammonia Ammonium Chloride Sal Ammoniac Ammonium Nitrate Ammonium Sulphate Calcium Carbide Carbide Calcium Carbonate Chalk Calcium Chloride Calcium Hydroxide Slaked Lime Calcium Oxide Quicklime Calcium Sulphate Gypsum Copper Sulphate Blue Vitriol Hydrogen Peroxide Peroxide Hydrochloric Acid Spirits of Salts Potassium Hydroxide Caustic Potash Potassium Nitrate Saltpetre, Nitre Sodium Bicarbonate Baking Soda Sodium Carbonate Washing Soda 2166 (NH,),SO, CaC, CaCQ, CaCl, Ca(OH), NaHCO, Na,CO, As ammonium hydroxide, a household bleach. Used to make chemical fertilizers, and in refrigerators. In Lechlanché cells, wet or dry, and in _ soldering fluxes. In explosives, and as fertilizer. A very important fertilizer. Used to make acetylene, formerly used in carbide lamps. Used in industrial processes etc. Used in building industry, and as drying agent in chemical processes. Used to make bleaching powder, plaster and in industrial chemical processes. Used to make slaked lime, cement, and as drying agent in laboratory. Used in manufacture of plaster of Paris. Used in copper plating, batteries, as a wood preservative and agricultural fungicide. Used as bleach and rocket fuel. Large number of industrial uses, in dye-stuffs and chloride manufacture. Drying agent, used in alkaline batteries and manufacture of soft soap. Used in manufacture of gun powder and fireworks, and as meat preservative. In baking powders, and health salts. Also has some industrial uses. Used as water softener and in manufacture of glass, soap powders etc.

Sodium Chloride Common Salt Sodium Hydroxide Caustic Soda Sodium Nitrate Chile Saltpetre Sodium Sulphate Glauber Salt Sulphuric Acid Oil of Vitriol Nitrous Oxide Laughing Gas Organic Compounds The Fuel Gases Coal Gas Producer Gas Water Gas Calor Gas Acetic Acid Vinegar Benzene Carbon Tetrachloride Chloroform Diethyl Ether Ether Ethyl! Alcohol Naphthalene Naphtha Octane Petrol Phenol Carbolic Acid Qe NaCl yee NaOH mixture CH,COOH CoH, ae, @ 3 cs e000 de: Ess] C,H,OH A ay Food seasoner, soap manufacture, production of chlorine, hydrochloric acid. Used in bleaching, dyeing, oil industry and in soap manufacture etc. Fertilizer. Used in manufacture of glass, paper, detergents and dyes. Very important chemical, starting point for explosives, fertilizers, etc, etc. Also used as drying agent. Used as anaesthetic by dentists etc. Used in industrial and domestic gas-heating appliances. Used in portable fuel units. Used as starting compound for acetates and other compounds. Also as organic solvent, and as vinegar. Solvent for fats, resins etc. Used in manufacture of drugs and dyes. Solvent used in industrial and dry-cleaning processes. Anaesthetic and organic solvent. Anaesthetic and organic solvent. Solvent used in manufacture of large number of organic compounds. In alcohol thermometers. Used as fire-lighting fuel, insecticide and starting point for manufacture of organic compounds. Fuel for vehicles. An antiseptic and starting compound for plastics, aspirin and other organic compounds. 2167

PROPERTIES OF MATTER Electromagnetism and the Velocity of Light FLECTRICAL charges lead a double existence. When they are stationary, in electrostatics, each charge exerts a powerful force on another neighbouring stationary charge. In current electricity the moving charges usually appear in droves, but, charge for charge, they exert far smaller forces on neighbouring moving charges. So electricity grew up in two vaguely related, but seemingly separate branches. Each developed its own set of units, the electrostatic units (e.5.u.) and the electromagnetic units (e.m.u.). Both sets of units are defined in a similar way. To relate them to quantities which can be measured in other branches of physics, they are defined by the forces charges exert on each other. Two identical spheres are similarly charged, and one centimetre apart. If they repel each other with a force of one dyne (just over a thousandth of the pull exerted on a mass of one gram by the Earth's gravity at its surface), then each sphere is carrying one e.s.u. of charge. Two wires carry current in the same direction, and are one centimetre apart. Because they set up a magnetic field, they exert equal and opposite forces on each other. The current flowing when the force between the The Dielectric Constant wires is one dyne is used to define the e.m.u. system of units, but here there are complications. The length of the wires matter, and so do the rates of flow of the charges. If the lengths of the wires are each one centimetre, then the rate of flow is one e.m.u. of charge per second. It is possible to write down equations that relate the force in dynes with the charges, in either e.s. or e.m. units. The equations are: Force (in dynes) = . first charge X second charge (e.s.) [distance between charges (in cm.)]? and Force (in dynes) = first charge (e.m.)/sec. cm. X second charge/sec. cm. [distance between charges (cm.)]" A force appears on the left hand sides of both equations, so both right hand sides must have identical dimensions, the dimensions of force. This is obviously impossible as the equations now stand, because the e.m.u. equation has an extra [length]? on the top and an extra [time]? on the bottom. Something is wrong with the equations, because they do not agree with each other dimensionally. Although the e.m.u. of charge is very much larger than the e.s.u. of charge, they are both charges, and must therefore have identical dimensions. The Magnetic Permeability PERMEABILITY =H CURRENT FORCE BETWEEN WIRES DEPENDS ON # The force acting between the charged plates is reduced by inserting a slab of dielectric material (an electrical insulator) in between them. The dielectric constant of air in the e.s:u. scale is |. A slab of iron increases the effect two neighbouring current-carrying wires have on each other. Iron has a higher magnetic permeability than air. Force is proportional to permeability. Because force depends on one-over-¢ in one equation and yp in the other, the quantity Tai appears in the dimensional equation. ei 2168 This discrepancy was a considerable stumbling-block, until it was found that the magic number 3 x 10! (30,000, 000,000) was the ratio of the values of e.m.u. and e.s.u. units of charge. 3 x 10? is the velocity of light, in centimetres per second. What is more velocity = length time is the extent of the discrepancy between the dimensions of an e.m.u. and the dimensions of an e.s.u. So, e.m.u.'s and e.s.u.'s are related in size and dimension by a factor equal to the velocity of light. This might seem rather a strange result, but in the late 19th century, Sir Arthur Ricker resolved the difficulty when he showed that one quantity had been missing from each of the 'force' equations. The dielectric constant, €, (practically 1 for air in e.s.u. but very high, for example, for water) had been missing from the electrostatic equation, and the magnetic permeability, wt, (also practically 1 for air in e.m.u. but very high for materials like iron) had been missing from the electromagnetic equation. When these were written in, it was found that the dimensions in the equations were put right provided the quantity te had the dimensions of a velocity. This ties up well with the electromagnetic theory of light. One of the consequences of this theory is that electromagnetic waves — like light waves, should travel through a substance with a velocity of 3 x 107° cm./sec. The Velocity of Light LIGHT SLOWS DOWN IN MEDIUM A consequence of the Electromagnetic Theory is that all electromagnetic waves (including radio waves and light waves) should travel through a substance with velocity Te The passage of an electromagnetic wave depends on the dielectric constant, and on the magnetic permeability.