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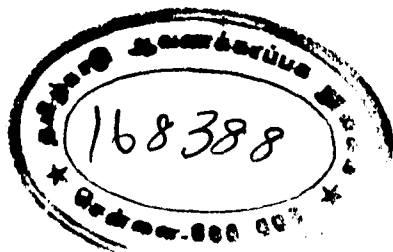
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GROWTH, REGENERATION, AND TUMOUR GROWTH

JOHN O. LAWS

MANY things from crystals to animal populations may grow, but as a biological term 'growth' indicates an increase in size of an organism in accordance with the regular plan which it has inherited from its ancestors. Regeneration is the replacement of parts which have been lost either by wear and tear or by accident. Tumour growth, which will be described in more detail later, is growth outside the normal pattern.

In many-celled organisms growth must be considered in relation to the cells which make up the body, and this process is in fact the result of three separate changes which are of varying importance in different phases of growth. The least important of these is increase in cell size. In general, cells of any one type in a given species of animal remain reasonably constant in size, and the only important exception is an increase seen when the second change is imminent, i.e. a division of the cell into two. Cell division is clearly a fundamental character of growth in many-celled animals since it provides the necessary bulk of cells to form the adult organism. Nevertheless one cannot say that the third type of cellular change, that of differentiation, is any less important, since without it the mere mass of cells would have little more character than the non-cellular slime moulds. Differentiation, the process by which cells take on their varied adult characters, involves the loss of most of the potentialities of the original germ cell, but at the same time the gain of function of one type or another. Even in the adult body some cells remain in a state of potential, rather than actual, function, often with active division (e.g. the basal cells of the skin and other epithelia), but as their ultimate function is in fact already fixed, they must be considered as differentiated.

The possession of such dividing cells is one aspect of the next attribute of the organism which we must consider – regeneration. Most organisms normally lose functional cells of their epithelia more or less continually during their lives and these are as continually replaced by division of the differentiated but non-functioning cells just mentioned; a process of physiological regeneration. In addition most species of higher animals possess to some degree the capacity to regenerate cells which are lost only by the accidents of life. This capacity varies from the complete reorganization of the body from a small portion, as in the flat worms called planaria, through the replacement of major parts containing many types of tissue, as in amphibian limb regeneration, to the much more limited capacity to regenerate only parts of an organ, as in mammalian liver regeneration.

All these types of regeneration have one thing in common; they reproduce more or less faithfully the lost structure and in no case produce excessive growth of any type. They are clearly therefore under some control similar to that seen in the development of the embryo and regenerative capacity is probably allied to the extent to which such a control exists in the adult animal. This expresses itself in the degree of de-differentiation seen during the process. In the planarian which is to rebuild three-quarters of the body, some cells at least must return to or have remained in a state in which they are capable of producing practically all types of adult cell (totipotent state). In the amphibian only certain tissues are present in the limb and the cell mass which is formed as a preliminary to regeneration will be able to produce a number of types of adult cell only (pluripotent state). In the mammalian liver it is probable that the various types of cell present in the regenerated organ are produced from similar cells in the remaining original liver which only de-differentiate in the sense of temporarily losing their function, without gaining any new potentialities for producing other types of cell.

It is interesting to compare the events in regeneration with what happens in tissue culture. If mammalian (or avian) cells are isolated and made to multiply *in vitro* (i.e. in a container, usually of glass, with suitable nutrients) they will, given suitable

conditions, do so indefinitely. In this they clearly differ from the cells of a regenerating organ which cease to divide when the normal bulk of the organ has been restored. It is not known whether this growth control functions as a positive inhibitor of growth in the intact animal, disappearing when regeneration becomes necessary, or as an active stimulus to growth appearing when injury occurs. The fact that cells *in vitro* divide only when stimulated by some active material, e.g. embryo extract, gives some support to the latter idea.

On the other hand, such cells in culture retain their differentiated characteristics, at least in potential form and, given the opportunity, will function again as, and only as, those originally isolated from the living organism. In this they resemble regenerating cells. In cancerous growth the cells behave similarly to those in culture, but with the fundamental difference that this occurs in the body itself. We see therefore that in considering the nature of cancerous growth, it is of considerable significance to study the process in relation to growth as a whole and to regeneration. The remainder of the present article attempts to present some aspects of this relationship.

A few years ago, in *Science News* 13, Professor E. C. Dodds stated that the mechanism of cancer growth and its method of killing its victims were understood perfectly. He was in this instance emphasizing that cancer in general obeys the same rules as other biological phenomena and possesses no mysterious 'influence'; nevertheless few workers on the problem today would agree with this statement in any more special sense. We in fact know little about normal growth mechanisms, let alone cancerous ones. And an increasing amount of evidence of the widespread biochemical effects produced by a tumour on the rest of the body which bears it, indicate that in fact it has an 'influence' which, while doubtless explicable, has not at the moment been explained.

In considering the origin of cancer it is customary to start with the mechanisms of carcinogenesis, physical, chemical, or unknown factors which apparently set in motion those changes which end up with the production of a cancer. The action of

these various factors is however very considerably influenced by the characters of the cells with which they come in contact. We will therefore begin here by reviewing some of the conditions which we know to affect the capability of a cell to become cancerous.

First one must consider the three categories of cells originally defined by Bizzozero many years ago. They are: (1) fixed cells which do not divide at all in the adult, for example nerve cells; (2) stable cells which do not normally divide under physiological conditions but may do so if like cells are damaged by some pathological process and have to be replaced, for example liver cells; and (3) labile cells which are continually lost and replaced under normal conditions, for example the cells of most epithelia. It is probably not coincidence that tumours occur most frequently in labile cells, sometimes in stable cells, in this case often following some other pathological process which has stimulated cell division in the organ, but never in fixed cells in the adult. Just what part in tumour production is played by cell division is not known, but some evidence which bears on this will be mentioned later.

Even in the case of labile cells, however, there are influences at work which further affect the liability to cancer production. An excellent example of this is the influence of hormones, that is of substances produced from the various endocrine glands including the sex glands and having a wide influence on many of the body tissues. At present it appears that this influence is more widespread than had been thought; for example it has been shown that an absence of hormones derived from the adrenal gland will prevent the formation of liver tumours in animals treated with certain chemical substances, although previously we had little evidence that these hormones had an action on the liver.

The most commonly considered example of the importance of hormones in tumour production, however, is the connexion of the sex hormones with cancer of the breast in the female and cancer of the prostate in the male. This connexion is probably a double one. In the first place the development of these organs

depends on the presence of the sex hormones at puberty, and clearly, if they do not develop then, no cancer is likely to arise later. Secondly, however, there is some other more subtle influence which is little understood but is inferred from indirect evidence – for example the greater occurrence of cancer of the breast in women who have children but do not suckle them than in those who suckle them normally.

This type of evidence has been strengthened recently by some which is more direct – concerning the part played by the pituitary gland in the production of certain experimental tumours. In the normal animal this gland produces hormones which stimulate ovarian activity. The ovary in turn produces its own hormones which, among other actions, depress the production of ovarian stimulating hormone by the pituitary, a type of self-regulating mechanism. If the ovary is transplanted from its normal site to the spleen of the same animal, it grows there and forms the normal hormones. As, however, the blood from the spleen passes through the liver before reaching the rest of the body (unlike the blood from the normal ovary) it is found that this latter organ destroys the hormones on the way through and none reach the pituitary. This gland therefore continues to pour out large quantities of ovarian stimulator which does reach the ovary through the blood going to the spleen. As a result the ovary continues in a state of abnormal activity and in many animals after some time tumours may appear in it.

Other influences than hormonal ones may apparently influence the appearance of tumours, but it is difficult to dissociate effects in the intact animal, and the action of such a factor as under-nutrition in reducing the incidence of tumours may be, in fact, an indirect one due to alterations in glandular action.

From what has been said of the greater occurrence of tumours in 'labile' cells it is clear that a potentiality at least for cell division is important in tumour formation. Whether actual cell division is necessary for this is not certain. The greater liability of 'stable' cells when actually stimulated to division leads to such a conclusion, but some recent experiments in which tumour formation took place in skin treated with carcinogenic

substance in spite of apparent suppression of all mitotic activity during the period of action by the use of cortisone suggest the contrary. However, until the action of cortisone is better understood, the exact significance of this result must remain in doubt, and it seems safe to say that, in most cases in which the process of carcinogenesis has been followed, it was associated with actual cell division in the tissue concerned.

The scene having thus been set, that is a number of suitably receptive cells being present, the first stage or 'induction' of the cancer can take place. The type of 'trigger' mechanism which sets the process off may be one of the many known physical or chemical stimuli, or, as in fact in most cases, unknown. Known stimuli include radiations, e.g. X-rays, and ultraviolet rays, and chemical substances of many types, e.g. naphthylamine which causes cancer in dye workers; tar and oils; complex hydrocarbons, either found in tar or synthetic, and many others. Some are known to be active in humans, some are active in a few species of animal and not in others; all produce cancers which are indistinguishable from those found occurring 'spontaneously' (i.e. without known cause) and it is because of this similarity that they can usefully be studied in the laboratory. We conclude that though the cancerous process can be initiated by diverse agencies, once started it follows the same path in each case.

Up to now no unique positive property of cancer cells has been found, but two aspects of their biochemical make-up are worthy of mention because, if not unique, they offer an interesting parallel with effects seen in regeneration. The first is the power of cancer cells to produce energy by aerobic glycolysis, a partial breakdown of glycogen, with a lessening but not absence of normal oxidative metabolism. A similar state of affairs is found in regenerative tissues of several species in the early stages when cell growth is most active but differentiation is not occurring. Possibly this mechanism represents an alternative route of producing energy when the normal oxidative mechanism is either insufficient, as in regeneration, or is diverted in some abnormal direction, as it is suggested may happen in the cancer cell.

The second biochemical effect, which has now been demonstrated in several ways, is the power of the cancer cell to alter the metabolism of distant tissues and to obtain its nutritive requirements at the expense of the host cells. A malignant tumour will continue to grow in a starved animal and gains weight at the direct expense of the host tissues, which waste. In regeneration in a number of species the phenomenon of 'protein flow' has been observed in which there is, during the early stages of regeneration, a loss of protein material from the rest of the body. Evidence tends to show that this is not an essential feature of the regeneration process, since artificial arrest of the flow does not apparently cause any slowing of the process; nevertheless it is a normal concomitant of regeneration, and suggests again that the cancer cell may show activities which, while absent in the normal adult cell, can nevertheless be found under special conditions of non-cancerous growth also.

Returning to the start of the cancerous process, it seems that the first step is an irreversible one. Whether, however, it affects all the cells exposed or only a few, we do not know. The reason for this is the existence of a latent period, often a long one in terms of the lifespan of the animal concerned, before the cancerous change becomes apparent. Either a few cells are rendered cancerous and then remain 'dormant' or 'latent' until stirred to further activity; or all or many cells are rendered 'potentially' cancerous but must be further activated to become true cancer cells which then at once show themselves. This difference in ideas is a real one, but need not concern us further here. What does emerge from this consideration is that cancer formation occurs in at least two stages, the 'induction' phase which we have discussed already and what is often termed the 'promotion' phase.

The existence of this latter may be inferred as above from the existence of a latent period in carcinogenesis, but in recent years it has also been demonstrated in several species that certain materials may act as promoting agents, speeding up the appearance of tumours which had been induced by true carcinogenic agents, although they by themselves do not produce

tumours. In general these agents do not influence the type of tumour produced or its ultimate fate, just its speed of development. Since continued treatment with a true carcinogen, although it speeds up tumour formation, also tends to alter the character of the tumours so produced, the separation of the two types of action is shown up clearly by these observations. The best known example of promoting, or as it is sometimes called 'co-carcinogenic', activity is that of croton oil on mouse skin.

Formation of a tumour is thus seen to proceed in stages, but this stepwise progress does not stop there. A tumour once formed may itself change in character, and this phenomenon has recently attracted attention under the name of 'progression'. The simplest form of this is probably seen in the change from 'benign' to 'malignant'. A benign tumour, like a wart, is characterized by excessive growth – increase in cell number with usually preservation of the function of the cells and no tendency to spread either by invasion of the local tissues or by extension to the more distant parts of the body. In most cases these characters persist throughout the life of the tumour, but occasionally it acquires the property of spreading and so progresses to the malignant phase. Malignant tumours may also arise directly without a benign phase intervening.

However, even when malignant the tumour may still progress further. It is common to find a considerable variation in appearance of tumours arising in one type of cell, and even in tumours of one organ, due to a variation in the degree of differentiated function that they show. Similar variation may occur between areas in one and the same tumour, or between the original tumour and its extensions in distant parts. Furthermore, if the tumour is made to survive through many generations by transplantation from one individual to another, such a change may occur in the course of this process leading to an entire alteration of the appearance of the cells of the tumour.

Not only is the external form of the cells subject to change but also characters of a purely biochemical nature. An example of this is seen in the progression of so-called 'dependent' tumours. These are tumours which can be produced in animals

which have been deprived of certain ductless glands and therefore of the hormones produced by these. The tumours which arise are invasive and can be transplanted from the original host to other suitable animals – provided that the new host also lacks the hormones which were missing in the original one. Later, if transplantation continues, the tumour progresses and loses this dependence so that it will then grow in a normal animal. A similar type of progression probably occurs in human cancers which are at first held in check by treatment, whether this be by hormones, chemicals of other sorts, or X-rays, but later grow in spite of continuation of the treatment.

The way in which such changes arise is not known, but it seems probable that they occur first in one cell and later come to predominate by a process of 'natural selection'. This is suggested by the way that they seem to show up particularly in circumstances in which cell survival becomes difficult, e.g. in spread to distant parts of the body, after transplantation and following treatments designed to kill as many cells as possible. Under such conditions selection has its greatest chance to operate.

Such are some of the changes which occur in cancer cells during the process of tumour formation and progression. They may be summarized firstly as changes in growth characteristics, some of which seem to parallel those seen in regeneration; secondly as loss of some of the functional characters of the normal adult cell; and thirdly as the acquisition of new ones, e.g. ability to spread to and survive in abnormal sites.

What then is the mechanism by which these changes are brought about? At present this is unknown, but several hypotheses have been put forward and will be considered here. Until just recently the most important modern theory was that of 'somatic mutation', and this is still probably the most favoured. It supposes that the carcinogenic stimulus, of whatever nature, causes a change in the genic make-up of a cell or cells which is then translated into active change in cell characters such as we have been discussing. Some exponents of the theory do not restrict themselves to one such change per cell but are willing to consider that six or seven may be necessary to provoke tumour

growth. In support of this theory much evidence has been adduced of visible changes in the chromosomal structure of cells of tumours or cells treated with chemical carcinogens or X-rays or 'radiomimetic' substances.* In fact similar changes can be seen in cells in tissue culture or in inflammatory conditions.

More recently attention has been drawn to the resemblance of tumour induction, development, and progression to epigenetic development in the embryo—albeit in an almost opposite direction, since in cancer cells progression is usually in the direction of greater growth freedom, autonomy, and loss of some specialized functions. Also one must emphasize that cancer cells even in very fast growing tumours which appear undifferentiated are in reality far from the embryonic cells which they may superficially resemble. Nevertheless the general characters of the change are somewhat similar to embryonic development and differentiation, and it may turn out to be profitable to consider whether the type of mechanism involved in the two processes may not be similar. Unfortunately the process of embryonic development and differentiation is itself little understood, but one fact which has emerged recently from the study of nucleo-cytoplasmic relationships in various organisms is that the cytoplasm plays more part in controlling cell make-up than some geneticists have been willing to admit. From this has emerged the idea that cell differentiation may be associated with the development of some balanced relationship between genic influence on the one hand and environmental influence transmitted through the cytoplasm on the other. Whether this involves the complete suppression of genic action and taking over of the appropriate functional direction of the cell by new cytoplasmic entities (cytogenes) or whether direct action of some genes continues with suppression in some way of the others is not known. It may be that some basic functions like growth remain under genic control while more specialized functions are directed by cytoplasmic centres. In any case they are all influenced by the environment and this detailed mechanism need not concern us further here.

*See, for example, the article by R. F. Homer in *Science News* 31.

It does seem, however, from a consideration of these points that the presence of some positive altered entity in the cancer cell which has so far been sought in vain is not a necessity in a hypothesis of cancer formation. We know that changes from growth to function and back occur in the type of cell we are considering in response to environmental demands and such switches are presumably mediated by some enzymic mechanism in the cell. It is possible to visualize a carcinogen affecting a dividing cell in such a way as to remove or inactivate an enzyme concerned in this mechanism so that a permanent deviation of energy to growth processes occurred. Such a deviation might of course be incomplete, leaving a cell capable of both dividing and functioning in an entirely irregular manner. In this connexion the recent suggestions by Professor M. M. Swann regarding an accumulation of energy for growth in a special form* are of considerable interest.

Such a scheme might account for growth abnormalities, but not necessarily for the capability of spread shown by the cancer cell. This is dealt with in the recent theory of Professor H. N. Green regarding loss of specific identity in cancer cells. He points out that normal cells must contain some marker substances which enable them to be 'recognized' in the body since such cells when misplaced in the adult are normally removed by scavenger cells. If, he suggests, these marker substances are lost, then the cell would be free to wander and invade other tissues without causing any reaction to remove them. In his theory he postulates that such a loss is due to a slight alteration being caused to these marker substances by a carcinogen, followed by antibody formation to the now 'foreign' material. Reaction of this antibody with the altered cell will normally cause cell death, but in a few cases the cell may survive by adaptive loss of the marker substance, leaving the cancer cell. Green has experimental evidence to support this immunological mechanism, but even without this, his conception of loss of specific identity appears to be a valuable one, and capable of direct experimental investigation.

*See *Science News* 34, 'Research Report'.

Thus it seems as if, from the biological point of view, there are now several lines of investigation open which may help to make clear the nature of the cancerous process. They appear to offer interesting links with the studies of biologists on normal growth and regeneration and it is to be hoped that a better realization of this may help to advance knowledge of all types of growth.

THE SCINTILLATION COUNTER

E. S. WILLIAMS

AT the base of the still widening applications of nuclear physics, useful and alarming, there lies a natural phenomenon discovered nearly sixty years ago—the phenomenon of radioactivity, in which certain atomic nuclei disintegrate spontaneously with the accompanying emission of energetic particles and radiations. The study of these unstable nuclei depends on the detection and identification of their disintegration products, so that the instruments used for this purpose are indeed the fundamental tools on which all else has been built. Of these instruments the Geiger counter has for a long time been the most important; but a new detector, the scintillation counter, based on different principles, is rising. In a few years it has been so far developed and applied that it is on the point of challenging the pre-eminence of the Geiger counter. It is the outcome of the harnessing together of a number of physical phenomena, which are of interest in themselves, but it also has a certain beauty of construction, a surprising versatility, and a fascinating history.

Sir William Crookes discovered in 1903 that when a zinc sulphide screen is exposed to α -rays in a dark room, the screen becomes brilliantly luminous. His apparatus consisted of the fluorescent screen, before which was placed a small point, capped with a trace of radium salt, and a simple microscope, through which the screen was viewed. This arrangement he called a 'spinthariscopes', and with it he demonstrated that the luminosity produced consisted of minute brilliant flashes of light which came and went with great rapidity. But the observations that he made with it were purely qualitative, and no measurements of any kind were made.

Crooke's discovery remained as a scientific curiosity, a 'beautiful experiment', until in 1908 E. Regener devised methods for counting the scintillations in order to determine the number

of α -particles falling on to the screen. The usual procedure was for two observers to view the screen, and for each to record on a moving tape the flashes seen. The number of coincidences on the tape represented the number of actual flashes, and elaborate precautions were evolved to eliminate so far as possible the observational errors. The scintillation technique became, in this form, the most widely used method in radioactive research, and numerous discoveries were made by its use, notably by Rutherford and his group.

One of the fundamental laws established in this period was the law of radioactive decay. This simple law states that the number of atoms n of a radioactive substance breaking down per second is proportional to the total number of atoms N present: that is, $n = \lambda N$, where λ is a constant characteristic of the substance. E. von Schweidler had shown in 1905 that this law could be deduced by assuming only that the disintegration of an atom was subject to the laws of chance and that the probability that a certain atom shall be transformed within a given interval of time is independent of the time which has elapsed since the formation of that atom. This probability is a constant which is the same for all atoms of the same kind. The law appeared to be adequate to account for the observed radioactive phenomena, but the experimental verification of the initial assumptions was the subject of research for a number of years, the scintillation technique being the most important tool in the work.

Scintillation counting held the field for about twenty years, although its versatility was limited by the method of recording the light flashes. Methods in which radiation was detected by means of the ionization it produced had from the first vied with those of scintillation. When the former methods culminated in the development of the proportional, and Geiger counters, these became the detectors of choice, and the scintillation counter fell into complete disuse for nearly twenty years. It is, however, worthy of note that even as late as 1932 important experiments were reported where the original two-observer scintillation apparatus had been employed.

There is a common factor, but also a difference, between these

two main types of detector. Both of them depend fundamentally on interaction between the radiation detected and the electrons of matter through which it passes. In both also there is a transfer of energy from the radiation to the material used for detection. In the scintillation counter, the energy absorbed is stored in so-called 'metastable states'. It is the reversion from the metastable to the stable state which results in the release, in the form of quanta, of the energy thus stored. These quanta are of energy within the visible or ultra-violet range. The possibility of producing 'metastable states' of suitable energy is a property of the material concerned. Those which possess it are known as 'phosphors' – substances which luminesce when exposed to energy-imparting radiation – and these may be either liquids or solids. In ionization methods, on the other hand, an electron is totally removed from an atom or molecule.

Now the wheel has turned full circle. Postwar improvements in methods of converting a light pulse into a pulse of current have led to the removal of the great limiting factor in the original scintillation apparatus, that imposed by the human eye. The replacement of the eye by electronic detectors of light has led to a gain of from 10 million to 100 million in the closeness of pulses which can be separated, depending on the type of detector used. This has been the primary innovation, although other developments have followed from it. In the following pages, the components of the modern scintillation counter will first be described, and then examples of its uses will be given.

COMPONENTS OF MODERN COUNTERS

In place of natural mineral crystals – usually zinc sulphide – there is now available a variety of phosphors suited to special uses. Between the phosphor and light-detector, there may, in some applications, be a 'light guide' so that the two former components can be physically separated. In place of the eye and recording tape, we have a photomultiplier, in which an initial small pulse of light is converted into a cascade of electrons, which is then further amplified, and finally the amplified pulses are counted. Figure 1 shows the arrangement schematically.

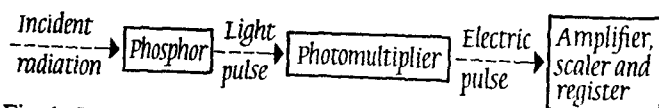


Fig. 1. To show the components in a scintillation counter. In some situations there may be a 'light guide' between phosphor and photomultiplier, and between the amplifier and register there are usually discriminating circuits.

Types of Phosphor

Certain inorganic crystals exhibit the phenomenon of luminescence, and these were always used in the days of the visual scintillation counter. More recently it has been shown that many organic crystals will scintillate when exposed to radiations and that this property occurs also in certain organic liquids and in polymerized (or plastic) solutions.

The luminescence of inorganic crystals is attributed to the presence of disturbances in the otherwise regular crystal lattice, caused either by a component of the lattice being missing or by the presence of an impurity. It has been found that the introduction into an otherwise pure crystal of a controlled amount of impurity gives a more efficient luminescence, the impurity being said to 'activate' the crystal. Common examples are zinc sulphide activated with manganese, and sodium iodide activated with thallium. A few crystals are efficient phosphors when in the pure state, as, for instance, the naturally-occurring scheelite (calcium tungstate).

These inorganic crystals are particularly useful as detectors of γ -rays. They contain atoms of high atomic number and are thus efficient in converting the energy of this highly penetrating radiation into energy which will be released in or near the visible band. This is particularly so with calcium tungstate where the tungstate atoms play a major role. In addition, these crystals can be grown to a large size (centimetre dimensions), and they are optically clear.

The luminescent process in organic compounds appears to be of a different nature. The phenomenon is here a molecular pro-

perty, and this accounts for the fact that luminescence occurs not only in organic crystals but also in liquids and vapours. The organic phosphors are all highly transparent and can often be grown into very large single crystals; the flash of light emitted by them is extremely short, the resolution time within the phosphor being of the order of 10^{-10} second.

The advent of luminescent liquids has made possible the almost unlimited increase in the size of phosphors, and this has led to a great extension of the field of application of scintillation counting. More surprising, perhaps, is the discovery that if an organic phosphor is dissolved in a material such as styrene monomer, and the solution is then polymerized, an efficient luminescent material is obtained. Such plastic phosphors are optically clear; they are not fragile; they can be obtained in large volumes; and they can be machined to any required shape. Examples of crystalline and plastic phosphors are shown in Inset 19.

Mounting the Phosphor

The selected phosphor has to be so arranged that the light it produces reaches the photomultiplier with as little loss as possible. Other things being equal, the more efficient the light collection the more sensitive will the complete instrument be as a radiation detector. When mounting the phosphor a number of requirements have to be met: (1) Light other than that produced within the phosphor has to be excluded; (2) the scintillations have to be efficiently directed on to the photocathode surface; (3) the phosphor must be protected from mechanical damage and, if hygroscopic, from moisture; (4) it must be possible to bring the phosphor into an advantageous position with respect to the source of the radiation to be detected; and (5) the phosphor should be screened from extraneous radiation.

Numerous types of phosphor-holder have been developed to meet these requirements, the details of each depending upon the purpose for which the particular counter was made. The problem of how to achieve maximum light-collection from the phosphor has taxed the ingenuity of numerous workers. The

idea of silvering the phosphor surface or of surrounding it by a silvered reflector was an early solution, but it has more recently been shown that in certain circumstances increased light collection results if all surfaces, except that through which light passes to the photomultiplier, are frosted, i.e. roughened to a matt finish.

If the phosphor is to be placed near to the source of radiation, it may be necessary to remove it some distance from the photomultiplier. This occurs when a scintillation counter is used in conjunction with the various types of particle accelerator employing huge magnets, since here it is best to mount the photomultiplier away from the region of intense magnetic fields. It also occurs when the instrument is required to investigate different parts of a wide X-ray beam where the rays would be scattered by the apparatus and lead to false results. In many cases it is necessary to measure the radiation inside a narrow cavity or opening where it would be quite impossible to introduce electronic equipment, a restriction which often applies in medical applications of scintillation counting.

In all these situations a light guide (or light pipe) is used. The possibility was beautifully demonstrated many years ago by Sir William Bragg when he showed how light could be transmitted along a curved water jet. The principle employed is that of total internal reflection; if light comes from inside a dense medium to the surface where it meets a less dense medium – e.g. glass/air – the ray will be reflected back into the dense medium provided the angle it makes with the perpendicular to the surface is greater than a certain maximum called the 'critical angle'. Thus if light enters one end of a rod of, say, 'perspex', it will pass along to the other end so long as any bends are sufficiently gradual for the light rays all to impinge upon the walls at angles greater than the critical angle. Where this requirement cannot be met it has been found to be satisfactory to silver the polished surfaces of the rod.

In its application to scintillation counting the light guide has taken a number of forms, and varies from a short probe a millimetre or two in diameter to one of two or three centimetres

diameter and several feet long. In all cases the crystal is cemented at one end of the guide and the photomultiplier at the other, the whole being surrounded by a light-proof outer covering. Some degree of success has been achieved also with flexible guides.

When it is necessary to investigate the radioactive content of small weak sources, or of sources emitting radiation of low energy, it is possible to introduce the source inside the crystal. An inorganic crystal is grown to a suitably large size and a hole of, perhaps, one or two centimetres diameter is bored into it to receive the radioactive samples. Such crystals provide an excellent device for the radioactive assay of liquids, but a simpler, and often adequate, method for this purpose is the provision of an annular bath surrounding a normal cylindrical crystal. Here the radioactive source is made to surround the crystal instead of vice versa. (Figure 2.)

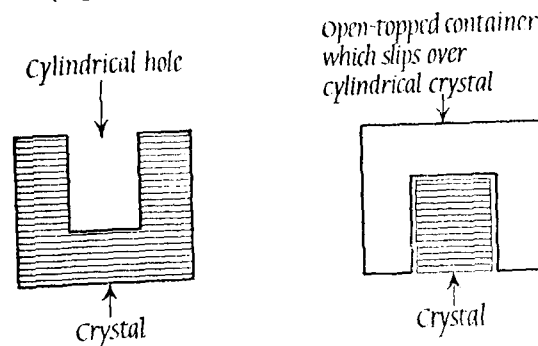


Fig. 2. Sectional diagrams showing two methods used to improve the degree of proximity of the source of radiation and the detecting phosphor. By such methods as these it is possible to increase greatly the efficiency of the detector and so to be able to assay very weak sources.

The assay of extremely low energy β -particles has often presented a difficult problem, and the situation becomes even more complex if the source is also a weak one. Recently, in the assay of certain biologically important materials, this problem was

solved by the neat method of dissolving the substance to be assayed in a liquid phosphor. In this case the 'geometry', as it is called, was so good that virtually 100 per cent of the emitted radiation was absorbed within the phosphor, giving a high detection efficiency.

The Photomultiplier

Whether directly or through a light guide, the light from the phosphor strikes a photosensitive surface in the photomultiplier and releases electrons from it. These electrons are accelerated by an electric field and impinge upon a second sensitive surface, where they give rise to the emission of a greatly increased number of electrons. These electrons are in turn accelerated in a similar manner and impinge upon a third surface, where the process is repeated. A number of such stages of emission and acceleration are included in the photomultiplier and at each stage the number of electrons is multiplied until the final, or collecting, electrode receives a very large pulse. This pulse is led into amplifying apparatus and thence to a register where it is recorded.

Two types of photomultiplier are in common use. An electrostatically focused type was developed by the Radio Corporation of America, and a linear, non-focused type was developed in this country by E.M.I. Laboratories, Ltd. The focused type (Figure 3) is fitted with a series of curved electrodes, or dynodes, each coated with a photosensitive material and so arranged that the electrons emitted at each dynode are focused on to the surface of the next in the series. In the linear type (Figure 4) the dynodes are composed of slats of metal arranged in venetian blind form; electrons released from the surfaces are dragged between the slats under the influence of an electric field and are accelerated to impinge upon the next dynode, multiplication being achieved as before. A fine-meshed grid is placed before each dynode to screen the secondary electrons from the influence of the electric field of the preceding stage, and so facilitates their escape. A photograph showing two examples of the linear type of photomultiplier is reproduced as Inset 20.

The Scintillation Counter

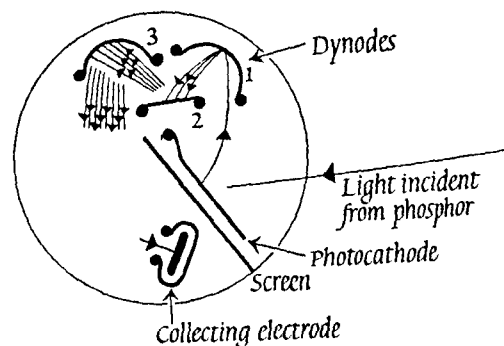


Fig. 3. The R.C.A. photomultiplier has curved dynodes so that the electrons are not only multiplied at each surface, but are also focused on to the surface of the next dynode in the series. Only the first three, and last, dynodes are represented.

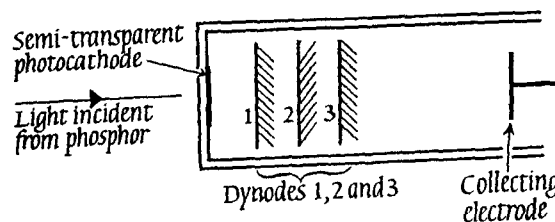


Fig. 4. The E.M.I. photomultiplier has dynodes in the form of venetian blinds. They are mounted in a line, the electrons in effect passing through each of them.

The use of electron focusing in the R.C.A. type makes it peculiarly sensitive to the influence of magnetic fields. Even the earth's field is capable of disturbing its function when the tube is placed in certain positions. The longitudinal, non-focused type is relatively insensitive to magnetic fields, and it has been found that with simple screening it is possible to operate such tubes satisfactorily in the vicinity of transformers and in other situations where stray magnetic fields are encountered.

In the focused type the electrons move in a uniform intense electric field and they have a short, and constant, time of transit from one dynode to the next. In the longitudinal type the secondary electrons are released into the space between the dynode slats, a region of low and unequal electric field intensity, and their transit time is long and somewhat variable. Since this variability occurs at every stage, the overall transit time of the electrons shows a considerable 'spread' and if one pulse follows closely upon another the two may merge and be registered as a single pulse. Such merged pulses are said to be 'unresolved', while those which are kept separate are said to be 'resolved' (Figure 5). The resolving time of a tube is the shortest interval between two successive pulses which are registered separately. That of the linear type is about 10^{-8} second (more than 10 million times better than the human eye), while the focused type has a resolving time more than 10 times shorter still, i.e. less than 10^{-9} second.

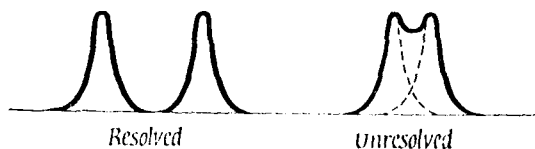


Fig. 5. Resolved pulses appear separately, unresolved ones are 'run together' and are counted as a single pulse.

Further Equipment

In addition to the phosphor and photomultiplier with or without an intermediate light guide, the complete scintillation counter requires certain items of standard electronic equipment, some of which are common to counters of other types. In the first place, a very stable constant voltage has to be supplied to the chain of resistors which divides the potential difference into the necessary steps to supply each dynode with its correct electric potential. This voltage is produced by a stabilized 'power pack'.

The pulse of current collected from the last dynode (the anode) is amplified before being passed to the recording device.

Because of the short resolution time of both phosphor and photomultiplier, the amplifier must have correspondingly 'fast' circuits. At one time it looked as if the more conventional electronic equipment might well set a limit to the performance of the instrument as a whole, but the recent development of transistor circuits solved the problem almost before it had arisen.

When counting the pulses caused by the incident radiation, it is necessary to exclude the many smaller pulses which arise from the spontaneous emission of electrons from the dynode surfaces. Similarly it is necessary that heavily ionizing radiation, if the counted in the presence of less heavily ionizing radiation, if the components of a heterogeneous beam are to be separated and the relative intensities of each component measured.

These requirements are met by passing the pulse from the amplifier through a discriminating circuit. This is an electronic device for sorting out the required size of pulse from all those arriving from the amplifier. It is as though each pulse were 'tried for size' before being passed on. The acceptable size can, of course, be varied at will by the user of the instrument; in many cases he can select either all pulses above a certain minimum or only those lying in the range between two given limits – the so-called 'channel' discriminator. (Figure 6.)

When the pulses have been sorted the selected ones pass to a

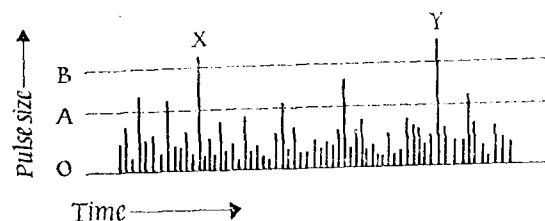


Fig. 6. Discrimination between different sizes of pulse. If the discriminator is set at level *A*, then all pulses of height less than *A* are ignored, and only those greater than *A* are counted. If a channel discriminator is used, then only those pulses falling, e.g., between *A* and *B* are counted, not only smaller ones being ignored but also those such as *X* and *Y* which are larger than level *B*.

scaling circuit where the 'powers of ten' are directed along different channels for recording, the total number of pulses being indicated visually or recorded graphically. Instead of this integral result it is quite easy to present the information as a 'count rate' so that the number of pulses arriving in unit time can be seen at a glance. Alternatively, in some applications, the result can be more conveniently expressed in the form of a picture on the screen of a cathode-ray tube. A complete scintillation counting unit is shown in Inset 21.

APPLICATIONS OF THE SCINTILLATION COUNTER

A scintillation counter is an assembly of apparatus the components and arrangement of which can be varied to suit the need. The technique as a whole is thus flexible. Its uses in research are of many kinds, but it will be possible to touch briefly on most of them.

Detection of Radiation

The mere detection of nuclear particles and radiations was the earliest application of the scintillation technique. It is still important, and some general notes on this use will be given first.

Alpha-particles can be detected with high efficiency with most of the known phosphors. A very thin screen of the luminescent material suitably mounted close to the window of the photomultiplier is robust and sensitive, and is also efficient for the detection of protons and deuterons. Since all these particles produce in the inorganic phosphors much greater flashes than do electrons or γ -rays, they can be distinguished from them and counted in a heavy background of β - or γ -radiation.

The organic phosphors are commonly preferred for the detection of β -particles. Here care has to be taken to ensure that the light shield surrounding the crystal is sufficiently thin to allow the β -particles to pass. In many cases it is preferable to give the β -particles free access to the phosphor, as is done in the detection of α -particles. In some instances, as mentioned earlier, it is even necessary to incorporate the source of β -particles within the phosphor itself.

Compared with other types of counter the scintillation counter has a very high efficiency for the detection of γ -rays. This makes it admirably suited for measurements on weak sources and in applications where the detector cannot be brought near to the radioactive material. Important fields are those of medicine and geology. In the former it is necessary to reduce a tracer dose to the smallest possible level, the ultimate limiting factor often being the efficiency of the detector. In geological applications the advent of this highly sensitive detector has made it possible to make aerial surveys for radioactive ores where before it was necessary to prospect on the ground.

The detection of neutrons has always presented a peculiar problem because they carry no charge and hence are not primarily ionizing. If the neutrons are 'slow', i.e. if they are of low kinetic energy, they are able to disintegrate lithium; thus crystals containing lithium are very sensitive to these particles. Lithium iodide activated with tin affords a method of detection that is nearly 100 per cent efficient. Fast neutrons release protons in the presence of hydrogen. This is an important property since all the organic phosphors are rich in hydrogen, and the entering neutrons will thus release protons which give rise to luminescence.

The detection of mesons and other unstable particles presents its own problems,* and numerous scintillation instruments have been designed for this purpose. Many of them are used for the study of cosmic rays, but recently the heavy particles produced by the larger accelerators have been investigated by similar methods.

Spectrometry

If the radiation to be investigated is heterogeneous, it is often of great importance to know over what band or bands of the energy spectrum the radiation spreads. It is also important to know what proportion of the total radiation possesses a particular energy. Such information throws light on the processes by which nuclear disintegrations take place, and it facilitates the detailed study of 'extraneous' radiation, such as the cosmic rays.

*See, for example, the article by R. M. Tennent in *Science News* 35.

This quantitative analysis of a beam of radiation is called spectrometry.

The spectrometry of the heavily ionizing particles such as α -rays, fission fragments, deuterons, and protons is simple when an inorganic phosphor is used. Here the efficiency with which the energy of the incident particle is converted into luminous photons is fairly constant for particles of the same kind, so that the intensity of the flash emitted is proportional to the energy of the particle. The various sizes of luminous pulse are sorted electronically, as already mentioned, and a graph is then plotted showing the number of particles in each range of energy.

The scintillation spectrometer has limitations when applied to β -particles. It has a poor resolution at low energies, but satisfactory results have been obtained at energies above 200 keV, and a number of novel devices have been designed to improve its performance. The proportional counter, on the other hand, has in turn been improved, and the direct scintillation spectrometer may here have little to offer except where the β -particles are to be measured in coincidence with γ -rays from the same source.

Less direct applications are, however, being reported. A typical example may interest readers of the article by Barnett in *Science News* 34. Here the β -rays of different energies were separated by means of a magnetic analyser similar in principle to the mass spectrometer, except that the extent to which the tracks of the particles were bent was here dependent on the velocities rather than on the masses of the particles. The detector was a phosphor which gave flashes proportional to the energy of the incident particles. Thus it was possible to select for counting only those flashes approximately corresponding to the correct energy. Scattered particles and other sources of spurious pulses were automatically ignored by a suitable setting of the discriminator circuit.

The spectrometry of X-rays and γ -rays poses a problem, since the radiation can interact in more than one way with matter through which it passes. For this reason, a mixture of γ -rays of comparatively few different energies may lead to quite a com-

plicated spectrum, often with overlapping peaks. The solution to the problem is simplified by scintillation techniques. This is especially so where the source is weak and where the component of interest happens to be only a minute fraction of the total radiation. Both calcium tungstate and thallium-activated sodium iodide give rise to sharp peaks in a graph relating counting rate to pulse height. The resolution in energy may be good enough in a simple case to enable the spectrum to be sorted out. If this fails, there is available a variant of the method in which the pulses caused by a single process of interaction are separated from those caused by others. The process selected is that in which a γ -ray photon collides with an electron in the manner of a billiards ball, so-called Compton scattering. The means for selection is provided by the short resolution in time which scintillation counting provides. By using two crystals, suitably placed, it is possible to count only 'coincidences' between flashes caused in the first crystal by the incoming γ -ray, and flashes caused in the second crystal by electrons ejected from the first. In that way, a spectrum is obtained in which there is only one peak for each energy of γ -rays.

The spectrometry of neutron beams, too, presents special problems. The tin-activated lithium-iodide crystal mentioned earlier as a detector gives poor results in spectrometry, but the scintillations produced by the recoil protons released in organic phosphors provide the basis of good-resolution neutron spectrometry. A number of neutron 'telescopes' using this principle have been developed for use with the large particle-accelerators. One of these, in use at Berkeley, California, employs four stilbene crystals in a line, the pulses from each being fed into triple and quadruple coincidence circuits before being recorded.

Very High Energy Particles

Early studies of particles of very high energy were confined to the investigation of cosmic rays, and gas counters and photographic emulsions were the detectors normally used. Coincident with the re-establishment of the scintillation counter as a research tool, the rapid development of the various particle-

accelerators took place, so that the main weight of research in this field swung from cosmic rays to supervoltage accelerators and from the older methods of detection to those using the scintillation technique.

The lines of research fall under two headings: scattering experiments, where the elastic and inelastic interactions between nuclear particles are studied, and meson experiments where the interactions of these short-lived, quantized 'pieces' of nuclear field are of interest. In scattering experiments observations are made to determine the absorption of the particles in matter, and to measure the relationship between the energy of the particles and the angle through which they are scattered in various absorbing materials. The study of mesons includes the measurement of the number produced as a function of the atomic number of the target atoms, the energy and type of meson obtained, and the degree of interaction of various mesons under different conditions.

Scattering experiments employ scintillation 'telescopes' (i.e. a number of phosphors in a line) of the type mentioned in connexion with the spectrometry of neutrons. The axis along which the phosphors are mounted can be rotated so that the particles arriving in any direction relative to that of the primary beam can be studied. The 'coincidence' technique previously mentioned is employed in the 'telescope' to enable the observer to identify the particle.

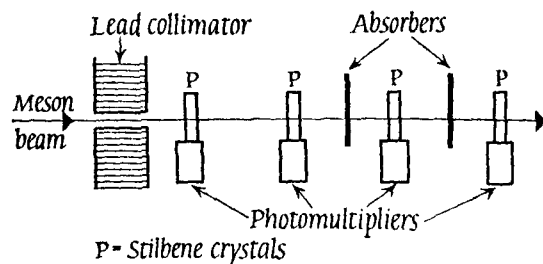


Fig. 7. An arrangement of a series of separate counters in a row is called a 'telescope'. Absorbers are inserted to adjust the effective range between detectors.

The very specific decay properties of mesons are used to identify them even though they are accompanied by an immense number of other particles. Apparatus of this class, which makes use of properties intrinsic to a particle, depends on the ability of the scintillation counter to respond specifically to events separated by extremely short intervals of time. It is in great contrast to the older equipment which measures only the kinematic properties. The proportional counter, for example, commonly used before 1948, often showed larger pulses caused by the background particles than were caused by the particles under investigation, and the unwanted particles were more numerous in the ratio of, perhaps, 100:1. The enormous advantage of the scintillation counter is thus obvious, but this is not all: the proportional counter has a long resolution time so that the maximum permissible counting rate is relatively low, the 'useful counts' able to be recorded thus being but a few per minute. With the scintillation counter, data which took several months to obtain can now be far more accurately recorded in a matter of days.

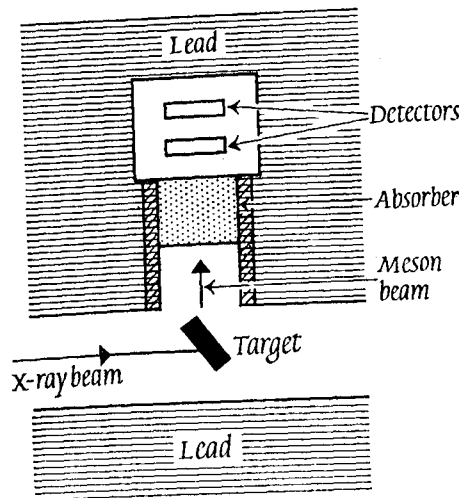


Fig. 8. Diagram of apparatus used to study the decay of π -mesons.

Suppose that X-rays of energy between 300 and 400 MeV fall upon a target. Then π -mesons will be emitted at approximately right angles to the direction of the rays and will enter a suitably placed detector. In one form of apparatus this consists of an aluminium absorber followed by the phosphor and photomultiplier of two separate counters (Figure 8). These are connected 'in coincidence'.

A positive π -meson decays into a μ -meson in about 2×10^{-8} second, and the μ -meson in turn decays into an electron in about 2×10^{-6} second. The entry of the π -meson into the first phosphor causes a pulse to be recorded, the emission of the μ -meson gives rise to a second pulse, and similarly the production of the electron in turn gives rise to a third pulse. These pulses are timed electronically and hence the process of meson decay can be studied. Absorbers can be placed between the phosphors to ensure that the events to be recorded take place where they are detectable, and a further phosphor and photomultiplier can be placed after the first two and connected in 'anticoincidence' with them so that a meson which passes right through without decaying will not be recorded.

Evidence for the production of a neutral π -meson by high energy X-rays from a synchrocyclotron was first obtained by the use of a scintillation technique. These neutral mesons decay into two quanta of average energy about 100 MeV, and, although the primary particles cannot be detected, these two γ -ray quanta produce scintillations by which their presence can be recorded. By an appropriate arrangement of the phosphors it has been shown that these quanta are emitted in pairs and that no other ionizing particles are simultaneously emitted. Measurements of the angular distribution provide data from which the approximate mass of the initial particle can be deduced. This has been shown to be about the same as the mass of the π -meson.

Nuclear Isomers

Many radioactive nuclei decay by β -emission, the positive charge on the nucleus being thus increased by one unit. The

nucleus so formed is of the element one place upwards in the Periodic Table. Since the lost charge carried with it a negligible mass the nuclear mass will remain unchanged, i.e. the 'mass number' is unchanged. Such a 'new' nucleus is called an isomer of the parent nucleus.

Many isomers are extremely unstable and decay with half-life periods as small as 10^{-9} second. That is, after one thousand-millionth of a second half of a given quantity of the substance will have disappeared. The decay of these rapidly disintegrating nuclei can be studied only by instruments capable of resolving such minute periods of time, and the scintillation counter employing organic phosphors is the only one that is feasible. The standard method employs two counters. One records the radiation accompanying the formation of the isomer, and the other records the radiation emitted in its decay. The electronic circuits of the second counter are so arranged that they will only pass a pulse to the register after a definite, though minute, time interval has elapsed since the passage of the pulse originating in the first counter. These 'delayed coincidences' are plotted against the delay time and hence the half-life of the isomer can be determined.

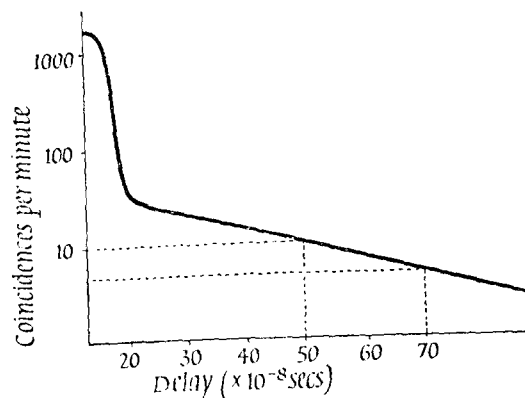


Fig. 9. See text.

The type of curve obtained is shown in Figure 9, where the coincidence rate is plotted on a logarithmic scale against the delay time. As the delay time is increased, the rate of coincidence on the graph is reduced linearly (i.e. the actual rate is falling according to an exponential law). In the hypothetical example shown the coincidence rate when the delay is 50×10^{-8} second is 10 per minute. The rate is reduced to 5 per minute when the delay is made 70×10^{-8} second, and thus in this case the half-life of the isomer is 20×10^{-8} second.

Large Volume Counters

In the section on types of phosphors, it was pointed out that the advent of the liquid organic phosphors made possible the use of extremely large scintillation counters. Recent work has been directed to exploiting this possibility, and it is likely that the large volume liquid scintillators will throw light on problems not approachable by other methods.

The spectrometry of very high energy electrons (greater than 1000 MeV), for example, requires massive expensive apparatus if magnetic methods are to be used. It is well known, however, that such electrons give rise to the formation of electron showers on passage through matter, the total energy of such a particle being finally degraded to that of numerous low-energy electrons. These can be observed in flight, through the small amount of (Cerenkov) radiation which they emit. Hence, if the shower is formed in transparent material, its energy can, in principle, be estimated. If the whole shower could be captured within a single phosphor, and the light collected in photomultipliers, then the size of the output pulses obtained would be proportional to the energy of the particle initiating the shower, and the spectrometry of such super-energy particles would be possible. In recent experiments designed to investigate the feasibility of this technique, scintillation probes were used within piles of plates of various absorbing materials, and the distribution of the energy within the shower, and the extent of the shower itself, were studied. The results indicated that large-volume Cerenkov counters seemed to have definite possibilities and it was hoped that

the spectrometry of electrons of energy as high as 10,000 MeV might ultimately be possible.

In another series of experiments a liquid phosphor of 300 litres volume and fitted with 90 photomultipliers was used. The primary purpose of this apparatus was to detect that most elusive of all nuclear particles, the neutrino. Such a large detector is, of course, also admirably suited to the detection of rare cosmic ray events and it has been used for this purpose. It must also have a very high γ -ray sensitivity and this fact has been turned to good account in measuring the natural radioactivity of human beings. Determinations can be made in a few minutes so that the counter offers a new method for the measurement of this, and other, very low activity sources.

Medical Applications

Wherever the properties of radioactive substances make them useful in the medical sciences, there will the many advantages of the scintillation counter be brought to bear on detection problems. The particular advantages of the instrument lie in its great γ -ray sensitivity and in the fact that it is possible to make the sensitive volume (i.e. the phosphor) very small, and to mount it at some distance from the associated electronic equipment.

The many applications of nuclear physics in medicine, and hence of the scintillation counter, may be placed in one of two groups: those where measurements are made clinically, and those where the measurements are part of laboratory investigations carried out on various body fluids and on biologically important substances. Perhaps the greatest impact of the scintillation counter in this field has been felt in the first group of applications, a number of tests being now possible which would have been inconvenient with the older forms of detector.

It is well-known that iodine is selectively concentrated in the thyroid gland. The rate at which this gland is capable, in a particular case, of removing iodine from the blood stream is a direct measure of its activity. In certain diagnostic tests a small quantity of the radioactive isotope of iodine is injected into a vein and the rate at which this accumulates in the thyroid gland is

measured by means of a detector placed over it. The advent of the highly sensitive scintillation counter has meant that much smaller quantities of the isotope can be used.

Similarly other isotopes are used in the investigation of the efficiency of the circulation in, for example, a limb. Either the isotope is injected into a vein at a remote site and the rate of build-up of radioactivity in the limb is observed, or it is injected at the site to be tested, and its rate of clearance therefrom by the circulation noted. In both of these cases the increased sensitivity of the newer detector leads to a great simplification in the experimental technique.

It has been mentioned in connexion with large-volume phosphors that a counter has been built capable of recording the natural radioactivity of the body. Such counters are useful in rapidly surveying for ingested radioactive compounds. The use of the older detectors would involve prohibitively long counting times on each individual. Even if these massive scintillators are not available in a particular laboratory, satisfactory 'total body content' measurements can be made with instruments incorporating the larger crystals of the inorganic materials mentioned earlier.

In a number of medical applications a detector having directional properties is required. This is achieved by surrounding the sensitive volume by a mass of material opaque to the rays to be detected: a suitable cylindrical hole in this material then effectively collimates the radiation and satisfactory directionality is achieved (Figure 10). For a given sensitivity a crystal is many times smaller than a Geiger counter so that the amount

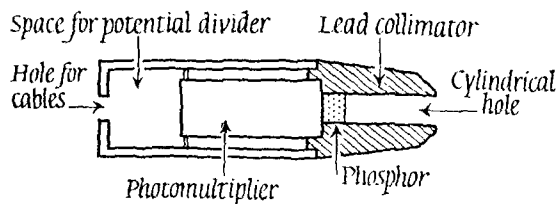


Fig. 10. Sectional diagram of a directional scintillation counter.

of collimating material needed is correspondingly reduced. This is an important point when it is realized that the common collimating material is lead and that most directional counters are required to be mobile.

When radioactive iodine is introduced into the thyroid gland via the blood stream, it is often of interest to determine whether it has been concentrated in one part of the gland more than another or whether all parts of the gland are equally active. This is particularly important when a lump is felt: is the mass functioning or not? The answer is given by means of a 'topographical survey' using a directional counter. The instrument scans the area to be investigated and records the radioactivity present at each point, the result being presented either as a picture on a cathode-ray screen, or as a recording on paper, or it may be necessary to observe the count rate at each point and record the results manually. Here the scintillation counter has, by virtue of its sensitivity, made the automatic recording of such surveys much easier, and the reduction in weight of the collimator has made the mechanical arrangements much more simple.

In the case of radioactive phosphorus the radiations emitted consist of β -particles only, and quite a small thickness of material is capable of absorbing them. Thus areas of phosphorus concentration cannot be located by detection of the β -particles unless the areas are on the surface of the body. The absorption process, however, gives rise to the emission of low energy γ -rays, and, if these are detectable, the presence of the β -emitter can be inferred. Again, the great sensitivity of the new counter has made this possible.

The fact that the phosphor can be placed at the end of a probe and the light piped back to the photomultiplier has led to the possibility of measuring the intensity of radiation within body cavities without the added complications of electrical leads, which the use of other detectors would involve. Since the phosphor can be quite minute, satisfactory probes for use in locating tumours in the brain have been made. Such tumours have been found, in many cases, to absorb some substances to a greater extent than the surrounding normal brain tissue, so that it is

only necessary to label a suitable substance with a radioactive isotope, administer it to the patient, and later locate the centre of radioactivity in the brain, thus assisting the neurosurgeon in locating the exact part to be removed.

In the laboratory the medical applications are intimately associated with the application of isotopic tracers to biochemical problems in general. The particular advantages of the scintillation form of detector are again due to its sensitivity, stability, and robustness, and as regards the assay of body fluids for radioactivity these advantages alone have resulted in a simplification of technique, a saving of time and labour, and an increase in accuracy. As an example the assay of radioactive iron in blood samples may be quoted. When the older methods of measurement were the only ones available, the blood had to be digested to free the iron, the iron then had to be extracted and electroplated on to discs, and each disc had to be assayed separately by a β -particle counter. It was only by this method that the minute traces of radioactive iron could be accurately assayed. Now the results which took some days to obtain can be accurately recorded in a matter of minutes without any preparation of the sample. A parallel case was that of the assay of plasma radioactive iodine where the plasma samples had to be carefully dried on discs before assay; now they are rapidly counted in a minute or two using a crystal with an annular sample chamber.

The detection of stray radiation, caused by inadequate screening of radioactive sources and by contamination, is of great importance in the protection of personnel from the harmful effects of nuclear radiation. Thus 'monitors', portable detecting instruments of as great a sensitivity as possible, are needed in every laboratory using radioactive materials. The many advantages of the scintillation counter have ensured that in this field, also, the instrument is widely recognized as an advance on older methods. In one instrument recently described a large scintillation screen was mounted at such a distance from a photomultiplier that flashes occurring at any point on it were received by the photocathode. The screen was provided with a rubber seal around the edges so that when it was pressed on a flat surface,

such as the floor or a bench top, stray light was excluded and any scintillations due to surface contamination were immediately detected.

CONCLUSION

This brief survey of the history, principles, construction, and uses of the scintillation counter will have given some indication of its importance in modern physics, and in the fields where physics is applied. Since its return to favour it has had a meteoric career, and is now one of the most valuable of all instruments. Only an outline sketch of the subject has been possible, and the following up of its many ramifications is left to the interest of the reader.

FURTHER READING

- Birks, J. B. *Scintillation Counters*. London, Pergamon Press, 1954.
 Curran, S. C. *Luminescence and the Scintillation Counter*. London, Butterworth, 1953.
Nucleonics, 12, No. 3 (March 1954).

These give a most useful survey, with numerous references to the already impressive volume of literature on the subject. To each of them, but especially to the first, the author made frequent reference while preparing this article.

THE SIZE OF ATOMIC NUCLEI

L. R. B. ELTON

ABOUT three years ago there appeared in *Science News* an article* which included a brief paragraph, headed 'Sizes of Nuclei'. It contained just about everything that at the time was definitely known on the subject. Since then, however, a great deal more information has come in of the most diverse character, so that, although much remains to be discovered, we are now much surer of what we do know than we were three years ago.

The first thing to be decided is whether a nucleus in fact has an exact size, and since nuclei are very nearly spherical, this means an exact radius. If nuclei really were like billiards balls, only smaller, this would not be difficult, but there seems no good reason why a nucleus should have a sharp boundary. It is much more likely that the density of nuclear matter decreases with increasing distance from the centre of the nucleus, so that the nucleus fades away rather than breaks off. This is, of course, not a problem peculiar to nuclear physics; a similar question arises if we inquire about the height of the earth's atmosphere. As a matter of fact, the density of the electrons which surround the nucleus of an atom does decrease in this manner, so that it is not possible to speak even approximately of the radius of an atom. On the other hand, in a nucleus the available evidence indicates that the density of nuclear matter remains roughly constant over most of the nuclear volume and then decreases rapidly to zero. To a first approximation it is, in fact, possible to work with an equivalent hypothetical nucleus of uniform density, a real billiards ball. (See Figure 1.) Many experiments actually cannot measure the detailed variation of the nuclear

charge density but give only a value of the radius R_u of the equivalent nucleus of uniform density.

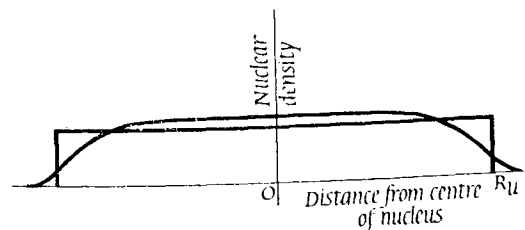


Fig. 1. Approximate variation of nuclear density with distance from the centre of the nucleus. The equivalent uniform density nucleus is also shown.

Even if a nucleus has then what may reasonably be called a radius, it does not follow that different methods of measuring this radius will yield identical results. Thus one method may measure the average radius R_u , another may measure the radius of the outermost limit of the density distribution. Fortunately the theory of the experiment often gives an indication which radius is being measured.

The methods of measurement can be divided into two groups according to whether nuclear or electric methods are used. In the former, the measuring probe is a light nuclear particle, such as a neutron, a proton, a deuteron or an α -particle, and this will interact with both the neutrons and the protons in the nucleus. Although some of these probes are electrically charged, their electric interaction with the nucleus is generally much smaller than their specifically nuclear interaction. In the latter, the probe is an electrically charged non-nuclear particle, such as an electron or a μ -meson, and this will interact only with the electrically charged protons, but not with the neutral neutrons. This method therefore can never give all the information about the nuclear density, but what information it can give, it is likely to give much more accurately and unequivocally; for we know so much more about electric forces than about nuclear ones.

*J. Little, 'Models of the Atomic Nucleus', *Science News* 23 (1953).

It had been established quite early in nuclear research that the density of nuclear matter varies very little from nucleus to nucleus. It then follows that the volume of a nucleus containing A particles—neutrons and protons—is proportional to A , and hence the radius to $A^{1/3}$. This fact is expressed by the law

$$R = r_0 A^{1/3} \times 10^{-13} \text{ centimetres} \quad (\text{Eqn 1})$$

where r_0 is a constant for all nuclei, although, as we have seen above, different methods of measurement are likely to give different values of this constant. It has been the purpose of recent experiments to confirm the above law and to establish the value of the constant r_0 .

NUCLEAR METHODS

Of the nuclear methods the most straightforward is undoubtedly that which employs neutrons. In this method a beam of fast neutrons is directed on to a sample of the material to be investigated, called the 'target', which partially absorbs and scatters the beam so that the outgoing beam is less intense. The reduction in intensity is due to the fact that some of the neutrons, instead of passing straight through between the nuclei, have interacted with them and have either been scattered by them or been absorbed into them. In the former case the neutron just changes direction, in the latter an unstable nucleus is formed, which sooner or later decays with the emission of one or more particles. In either case the neutron has been removed from the beam.

Now if a beam of neutrons acted like bullets from a machine gun, it would be easy to calculate the radius of the nuclei in the target. Suppose that a fraction n of the incident beam has been removed by the target, and suppose that the target has N atoms per square centimetre. If the radius of the nucleus is R , then each nucleus presents an effective area πR^2 to the incoming beam, and if the cross-sectional area of the beam is X , we have

$$n = NX\pi R^2 / X = N\pi R^2. \quad (\text{Eqn 2})$$

As N is known and n can be measured, R can be found.

I have said nothing so far about which radius is measured by

this method. It is clear that the incident neutron will be affected by a nucleus in the target if it comes within the nuclear force range of any neutron or proton. The range of nuclear forces is admittedly short, of the order 2×10^{-13} centimetres; nevertheless the radius measured by this method should be larger by roughly this amount than the nuclear radius proper, and this latter is the radius of the outermost limit of the distribution. We would expect then to find a law of the type

$$R = r_0 A^{1/3} + b, \quad (\text{Eqn 3})$$

where b should be approximately 2×10^{-13} centimetres. This represents the additional range of the nuclear forces.

Unfortunately, however, our picture is grossly oversimplified. Neutrons are not bullets. It is one of the most fundamental facts of modern physics that all particles also have wave aspects. In consequence neutrons which, if they were purely particles, would miss a nucleus, may be deflected by diffraction, in the same way in which a light wave is diffracted round an obstacle.

Thus the effective area which each nucleus presents to an incoming neutron is much greater than πR^2 . It varies in a complicated manner with the energy of the neutron, being much greater for smaller energies and decreasing to $2\pi R^2$ for very large energies. To be able to interpret the experiments simply, it is therefore necessary to use very fast neutrons, of energies greater than 50 MeV.* But at those energies the nuclei themselves are partially transparent to the incident neutrons, so that a neutron may pass right through a nucleus without being deflected. This reduces the effective area of the nucleus by an unknown amount. The most reliable results have therefore been obtained at somewhat lower energies, mainly at 14 and 25 MeV. The most complete are those of J. H. Coon, E. R. Graves and H. H. Barschall of the University of California. When Equation 2 has been corrected for diffraction effects, it can be written

$$R = \sqrt{n/cN\pi}, \quad (\text{Eqn 4})$$

*1 MeV is the energy acquired by an electron when accelerated through a potential difference of 1 million volts.

where c is a number that depends on the energy of the incident neutrons and decreases towards 2 for very large energies. The dependence of c on the energy has been evaluated under certain simplifying assumptions by V. F. Weisskopf and H. Feshbach of the Massachusetts Institute of Technology.

If it is correct for all nuclei that the radius is proportional to the cube root of the volume (Equation 1), then the dependence of R on A is given by Equation 3, and we can readily arrive at the constants b and r_0 of the latter equation. A plot of R against $A^{1/3}$ should give a straight line with intercept b on the y-axis and

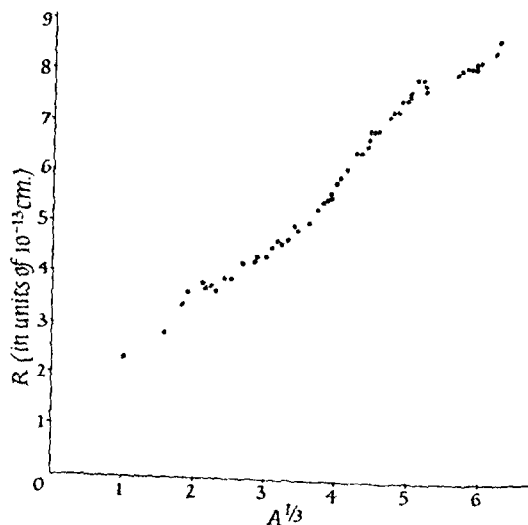


Fig. 2. A plot of $\sqrt{(n/cN\pi)}=R$ against $A^{1/3}$. For explanation see text. (Experimental results by J. H. Coon, E. R. Graves, and H. H. Barschall, *Physical Review*, 88, 564, 1952.)

gradient r_0 . The experimental results have been plotted in Figure 2. It is clear that the points lie roughly on a straight line, so that the assumption contained in Equation 1 is very approximately correct, but that there is a considerable latitude in the fitting of

the line. One can say from the graph that $r_0 = 1.4 \pm 0.1$ and that the value of b is around 1×10^{-13} centimetres.

Similar experiments using charged particles as probes are much more difficult to interpret, because of the electric interaction which is added to the nuclear one. Thus D. M. Chase and F. Rohrlich of Princeton University in interpreting the results of the scattering of 18.6 MeV protons obtain $r_0 = 1.42$, and G. W. Farwell and H. E. Wagner of the University of Washington find that their experiments on the scattering of α -particles of energies varying from 13 to 42 MeV yield $r_0 = 1.5$. Other results range as high as $r_0 = 1.8$. The uncertainties in these results are generally not due to experimental inaccuracies but to the difficulties that arise and the approximations that have to be made in applying mathematical theory to the interpretation of the experimental results. A further uncertainty arises from our very incomplete knowledge of nuclear forces.

A quite different method is based on the type of radioactive breakdown in which α -particles are emitted (α -decay). To understand this we have to introduce the idea of potential energy. This is the energy which a system possesses in virtue of its position and is equal to the work done in bringing the system together from infinite separation of its constituents. It is clear, then, that if the parts of the system repel each other, the potential energy of the system increases as we bring them together, and if they attract each other, it decreases. Now in α -decay we are concerned with an α -particle and the residual nucleus which remains after the decay. Both these are positively charged and so repel each other. If then we bring them together from infinite separation, the potential energy increases until the point is reached at which the nuclear forces take over. These are strongly attractive, and so the potential energy decreases rapidly. Once the α -particle is inside the nucleus, it is attracted equally from all sides, so that there is no resultant force. The potential energy is then constant. All this is illustrated in Figure 3.

Now the total energy E of the α -particle is the sum of its kinetic energy and its potential energy. This energy has been drawn in on Figure 3. According to classical physics, an α -particle

inside the nucleus is like a ball-bearing rolling in a saucer. Either the ball-bearing has enough kinetic energy to go over the rim of the saucer, in which case it will do so at once, or it has not, in which case it will roll about the saucer for ever. (We neglect friction in our analogy.) The energy E of the α -particle is certainly as drawn in Figure 3, i.e. it is insufficient to surmount the barrier. However, according to the laws of wave mechanics such a particle has a small though definite chance of 'tunnelling' through the barrier and appearing on the other side. This is the

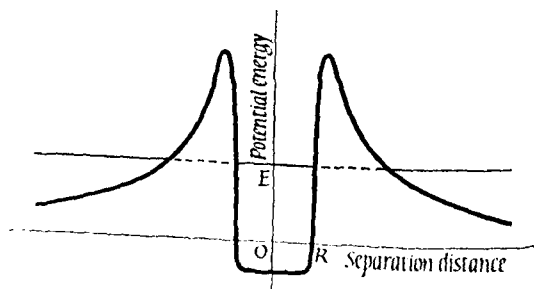


Fig. 3. The potential energy of a radioactive nucleus, plotted against the separation distance r of the α -particle and the residual nucleus. The energy level of the α -particle is also shown.

explanation of the phenomenon of radioactive decay. The magnitude of this chance depends very critically on the length of the tunnel, and as the exit of the tunnel is very well known (it only depends on the electrostatic forces) we have here a measure of the nuclear radius. Which particular radius we are measuring is, however, not clear. This method can only be used for radioactive nuclei, for all of which A is greater than 208. The values of the radius found fit in quite well with the values required for heavy nuclei from neutron scattering.

ELECTRIC METHODS

We next turn our attention to electric methods. By far the most important of these is the scattering of very fast electrons by

atomic nuclei. This is basically equivalent to looking at the nuclei through a microscope. Now the resolving power of a microscope is limited by the wavelength of the light used, i.e. it is impossible to distinguish detail which is finer than the wavelength. The only way to obtain wavelengths as short as nuclear dimensions (a few times 10^{-13} cm.) is to use high energy electrons. The wavelength (λ) of an electron is most conveniently expressed in terms of the so-called Compton wavelength (λ_0) – which for the present purpose can be regarded merely as a reference wavelength* – and the total energy, E , of the electron. The latter is made up of the rest-mass energy of the electron, mc^2 , and of its kinetic energy. It turns out that, if E is measured in MeV, we can write (very nearly)

$$\lambda = \lambda_0 / \sqrt{(4E^2 - 1)}. \quad (\text{Eqn 5})$$

We can then construct the following table:

E (MeV)	1	20	50	100	200
λ (10^{-13} cm.)	220	10	4	2	1

Radii of nuclei vary from about 3 to 8×10^{-13} centimetres. Thus 1-MeV electrons will 'see' a nucleus as a point, 20-MeV electrons will detect that a nucleus has a finite size, and electrons of 100 to 200 MeV energy will probe the detailed structure of the electric charge distribution within the nucleus. All these predictions have been fully verified. In particular, the very accurate experiments of R. Hofstadter and co-workers of Stanford University with electrons of energies ranging from 84 MeV to 183 MeV have for the first time given us direct information about the structure of the nucleus. The electrons are accelerated by a linear accelerator,† and it is planned to extend the accelerator so as eventually to give energies at least twice as large.

The nuclear radius which is measured by all electric methods is the equivalent uniform density radius R_u . Let us try to see the reason for this difference between the nuclear and the electro-

*In fact, $\lambda_0 = 386 \times 10^{-13}$ centimetres.

†For information on linear accelerators see the article by Howard-Flanders and Haslett in *Science News* 32.

magnetic results. It lies essentially in the different character of the nuclear and electric forces. Although the nuclear forces are very strong, their range, as has been pointed out, is exceedingly short, of order 2×10^{-13} centimetres, which is only a fraction of the nuclear radius. It was because of this fact that we came to the conclusion that nuclear methods measured the radius of the outermost limit of the distribution. The range of electric forces on the other hand is practically infinite, and certainly extends to distances hundreds of thousands of times as great as the range of nuclear forces.* In consequence the effect on the electrons is an average one, due to all the protons in the nucleus, and hence the radius measured is also an average one.

The experiments yield a value $r_0 = 1.20 \pm 0.05$, and possible charge distributions which will fit the experimental results are shown in Figure 4. The distribution with the dip at the centre

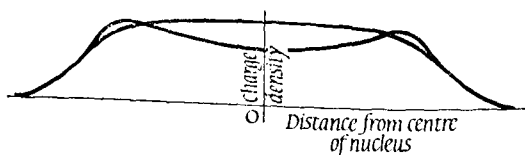


Fig. 4. Two possible distributions of the electric charge in the nucleus.

gives a slightly better fit to the data, although not sufficiently better to make it possible to distinguish definitely between the two possibilities. It is actually quite likely that because of the electrostatic repulsion between the protons, these will tend to push each other outwards, leaving a lower density of charge at the centre.

Another method depends on the so-called isotope shift. Isotopes are atoms with the same number of protons but a different number of neutrons. They are chemically indistinguishable, and

*The electrons in the outer atom are certainly influenced by the electric forces of the nucleus, and the linear dimensions of an atom are of the order of 10^{-8} centimetres.

therefore normally occur as a mixture. Now the energies of the electrons in an atom are determined by the nuclear charge distribution, and if we consider two atoms whose nuclei each have the same number of protons, but one has one or two more neutrons than the other, then the corresponding electron energies will differ very slightly. The difference, which shows up as a splitting of a spectral line, is exceedingly small, but being measured by spectroscopic methods – than which there are probably no more accurate methods in physics – it is just detectable. From this tiny energy difference it is possible to obtain a measure of the increase in nuclear volume caused by the presence of the extra neutrons, and hence of the nuclear radius. Most of the experiments have been done by P. Brix and H. Kopfermann in Göttingen. These, too, can be fitted very well with $r_0 = 1.2$.

An interesting method that has been developed recently depends on the use of μ -mesons. These particles are very similar in many ways to electrons, except that they are about 200 times heavier. If they are captured by atoms, they perform orbits, just as electrons do, except that because of their greater mass the radii of these orbits are 200 times smaller. This brings them very close to the surface of the nucleus, so that the energy of a μ -meson in an orbit depends very sensitively on the nuclear radius. The effect is much greater than that of the isotope shift, but unfortunately the accuracy of measurement is much less since μ -mesonic atoms are not easy to produce, and are extremely short-lived. In the same way that ordinary atoms emit light of definite wavelengths, a μ -mesonic atom emits X-rays, and the wavelength of the X-rays determines the nuclear radius. Experiments along these lines have been performed by V. L. Fitch and J. Rainwater of Columbia University, and these give values of r_0 close to 1.2.

Lastly we can obtain a value of r_0 from the amount of energy stored in a nucleus. This quantity was referred to as E by Little in his article. It contained, as he pointed out, a term due to the repulsion energy of the protons, which is a form of potential energy. The electrostatic potential energy between two protons,

which are distance r apart, is well known to be e^2/r . If we then consider the potential energy of each of the Z protons of a nucleus in the field of the other $Z-1$, their total potential energy will be directly proportional to $Z(Z-1)e^2$ and inversely proportional to their mean distance apart, which is of the order R_0 . An exact calculation shows that the electrostatic potential energy term is actually given by

$$3Z(Z-1)e^2/5R_0.$$

Clearly $3/(5R_0)$ is the quantity called d by Little.* It is found by fitting the complete formula for E to the experimental values for the stable nuclei, and the resulting value of r_0 is $r_0 = 1.22$.

MIRROR NUCLEI

There is one more piece of experimental evidence, and this at one time was considered to give the most reliable value of R_0 . Although it is now somewhat discredited, it must be mentioned here, since it held the field for so long.

Among the light nuclei there are pairs which consist of the same odd number, say $2M+1$, of particles, but one of them has M protons and $M+1$ neutrons, while the other has $M+1$ protons and M neutrons. A typical example is the pair N^{14} and O^{14} . Such nuclei are called mirror nuclei, and of two mirror nuclei one is always β -radioactive and decays into the other. This will now be shown.

As far as the first M protons and M neutrons are concerned, two mirror nuclei are quite identical. But if the last particle is a proton, it is repelled electrostatically by the other protons and is therefore much less firmly bound to the nucleus than a corresponding last neutron. For that reason the proton-richer of the two mirror nuclei decays into the neutron-richer, the last proton decaying into a neutron with the emission of a positron. The energy given out in this radioactive decay must just be equal to the electrostatic potential energy that the last proton had in the field of the other protons before the decay. On the assumption

*Little replaced $Z(Z-1)$ by Z^2 . The difference is not significant.

that the chance of finding this last proton at a particular place in the nucleus is the same as that for any other proton in the nucleus, the potential energy referred to above can be evaluated and turns out to be

$$E = 6Me^2/5R_0. \quad (\text{Eqn } 6)$$

The energy E can be determined experimentally, and a plot of E against $A^{-1/3}$ should then be a straight line, from the gradient of which we can find r_0 .

It is certainly found that the experimental points lie on a straight line, in fact the accuracy with which they lie on this line is quite remarkable; but the value of r_0 thus found is 1.46. As this is the result of an electric measurement, it must be compared with the values obtained by other electric methods. The discrepancy is considerable. What has gone wrong?

Although the complete answer to this question is rather complicated, a partial answer can be given at once. The assumption made earlier about the chance of finding the last proton at a given place in the nucleus is almost certainly wrong. Just because this is the one proton that decays, it is the one that is the least firmly bound of all the protons in the nucleus, while our previous assumption was equivalent to assuming that it was just as firmly bound as all the other protons. If account is taken of this, then the nuclear radius must certainly come out smaller than before. Recently a much more exact and fundamental approach by D. C. Peaslee of Columbia University has, in fact, shown that a value $r_0 = 1.2$ is at least not in conflict with the evidence from mirror nuclei.

THEORETICAL TREATMENT

An interesting theoretical confirmation of the experimental results has recently been given by M. H. Johnson and E. Teller of the University of California. In the same way in which we described the α -particle as being held in a potential energy well (see Figure 3), we can also describe neutrons and protons being held in a nucleus. The only difference is that we are now dealing with a stable nucleus. The total energy of a neutron or a proton

in such a nucleus must therefore be less than the corresponding energy at infinite separation. Only in this way can there be no tunnelling through the barrier.

Since a neutron is uncharged, it is not affected by a nucleus until it comes within range of the nuclear forces. Its potential energy curve is therefore as in Figure 5a. For protons we have to add to this the electrostatic potential energy; this is positive at all distances, because both the proton and the rest of the nucleus are positively charged and so repel each other. The electrostatic potential energy of the proton is shown in Figure 5b and the total potential energy of the proton in Figure 5c, which is obtained by adding the two previous curves.

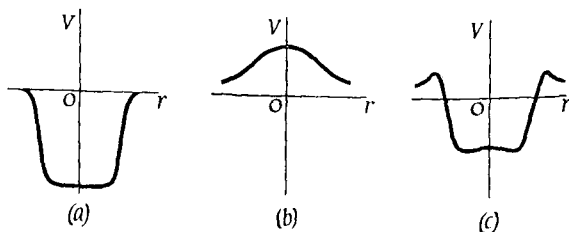


Fig. 5. The potential energy V as a function of the distance r from the centre of the nucleus for (a) a neutron in a nucleus, (c) a proton in a nucleus. In (b) the purely electrostatic part of the potential energy of a proton in a nucleus is given.

It is now a consequence of wave mechanics that there are only certain energies which a particle held in such a well can have* and that there can only be one particle with any given energy in the well. If we imagine that we are filling up such a well, then the first particle goes into the lowest energy level, the second into the next, and so on. Because of the sloping walls of the well,

*Mathematically this is equivalent to the fact that, for instance, an organ pipe can only emit certain notes, the fundamental and the overtones, but no others.

particles with higher energy can spread further outwards than those with lower energy.

To complete the argument we proceed to prove that the most energetic proton and the most energetic neutron in a stable nucleus must have very nearly the same energy. This is because of the phenomenon of β -radioactivity, by which a proton can

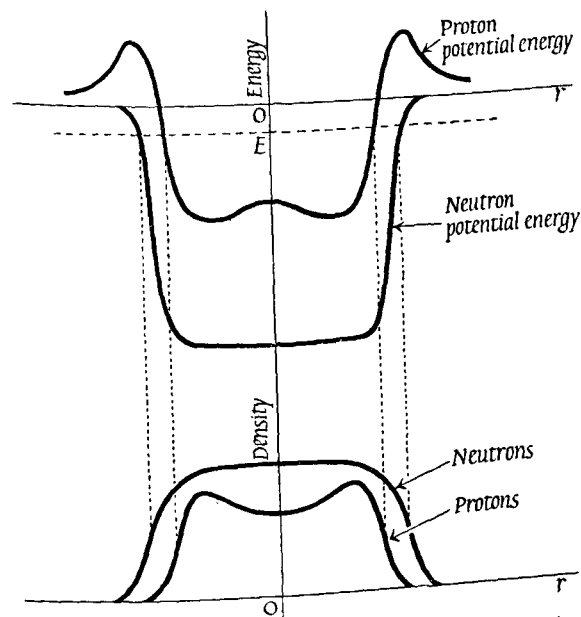


Fig. 6. Potential energies and densities for protons and neutrons inside a nucleus. (Redrawn from M. H. Johnson and E. Teller, *Physical Review*, 93, 357, 1954.)

change into a neutron with the emission of a positron, and a neutron into a proton with the emission of an electron, if such a change is energetically possible. In practice this means that if, say, the most energetic proton is more energetic than the most

energetic neutron it will change into a neutron with the emission of a positron and some energy, and vice versa. As a stable nucleus by definition is not β -radioactive, the most energetic proton and neutron in a stable nucleus must have the same energy.*

In Figure 6 the potential energy curves of the neutrons and protons have been drawn on the same diagram, together with the energy level E of the most energetic proton and neutron. From this it is possible to deduce the variation of the proton and neutron densities. This is done in the lower half of the figure.

We see at once that (a) there are more neutrons than protons in a nucleus, (b) the neutron density goes further out than the proton density, and (c) the proton density has a dip in the middle. All these facts are borne out by experiment. It should be noticed that we have here a slightly different reason for expecting nuclear methods to give a larger radius than electric ones. The nuclear methods measure the outermost limit of the neutron distribution. This is important, because the mere difference between the mean radius and the radius of the outermost limit of the same distribution is unlikely to be of the order required by experiment.

CONCLUSION

On the whole, as we have seen, theory and experiment are in very good agreement, and, although there are discrepancies, there are no contradictions. There is, however, a striking contrast between the results obtained by nuclear and those obtained by electric methods. This contrast does not so much lie in the discrepancy in size, which has been explained, but in the difference in the degree of agreement obtained. The agreement among the electric methods is almost suspiciously good; one is not used to perfect agreement in nuclear physics. It is due, of course, to the fact that there is little we do not know about electric forces. The vagueness of the results obtained by nuclear methods re-

*More detailed considerations show that a very small energy difference between the most energetic proton and neutron is permissible.

fects our ignorance of nuclear forces, and until we know much more about them generally little further progress can be expected in this specialized field.

Note added in proof. Recently Feshbach and Weisskopf have applied a modified theory to Barschall's results and have obtained very good agreement with $r=1.45$. (See H. H. Barschall, *American Journal of Physics*, **22**, 517, 1954.)

SULPHUR IN PICTURES

SULPHUR is one of the twelve most abundant chemical elements; there is more of it in the earth's crust than, for example, of aluminium, chlorine, or calcium, none of which are regarded as scarce. It is also one of the few elements, apart from the atmospheric gases, which are found on an appreciable scale as the free element. Further, since it melts at $113-115^{\circ}\text{C}$, it can be extracted, when overlain by a dome of pressure-resistant rock, by using a piped supply of superheated water to melt the sulphur, and air under pressure to drive it to the surface; this is the Frasch method, introduced in 1891. But for these two facts, it is unlikely that there would at any time have been talk of a shortage of sulphur; the ease with which the limited, suitable supplies can be exploited in this way has been too great.

Most of this 'easy' sulphur comes from Texas and Louisiana in the United States, and, for reasons not wholly connected with normal supply and consumption, it was decided in 1951 by that country that both internal consumption and export should be restricted. Although there was nothing basically new in the situation, the result was a sulphur panic which in greater or less degree affected most countries. The reason for this was the heavy dependence of the chemical industry on sulphuric acid, coupled with the need of other industries, notably rayon, for sulphur as such. Figure 1 shows the course of sulphuric acid production in Great Britain from 1925 onwards; apart from the erratic movements of 1925 to 1933, the curve corresponds with a growth-rate at about 4 per cent compound interest. The accompanying photographs, to which these notes are a supplement, illustrate the course of events since 1951.

Inset 1 shows a museum specimen of native sulphur crystals. The fact that the specimen comes from Sicily is a reminder that this island was the great historic source of sulphur. Except for local consumption, the mining of sulphur by other than the Frasch method has suffered by comparison during more than a

generation. As well as Sicily and Italy, rock sulphur is worked in a number of countries. The most important sources are deposits at high altitudes in the Andes, and others in Japan. Production is, in general, well below capacity. The future use of these deposits probably depends mainly on improvements in methods of extraction, on which research and development are proceeding. There are several claims of ability to produce in competition with Frasch-method sulphur, and these are about to be put to practical test.

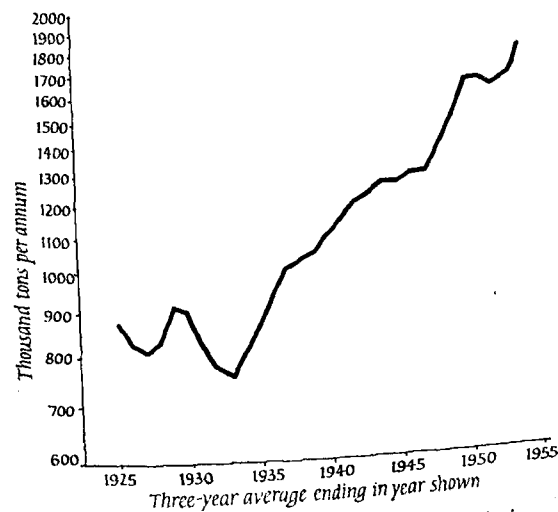


Fig. 1. Sulphuric acid production in Great Britain.

The following group of photographs (Inset 2-5) illustrates developments in the production of sulphur by the Frasch method. The most recent of these developments has been the addition of production from Mexican sulphur comes to that from the United States. In round figures, Mexican output now adds about 10 per cent to that in the U.S., and will prolong supplies. In the same period, further and more important

discoveries were made in Louisiana. These reached full production early last year, and have restored the American position. Reserves are variously estimated to cover the present rate of production during a period of about 10 to 20 years. In practice, tailing off may be expected, which may be made good by Mexico. An off-shore sulphur dome in the Gulf of Mexico is now also being examined.

Elemental sulphur extracted by the Frasch method is expected to account for some 43 per cent of total production in 1956, with a further about 3 per cent contributed by other native sulphur. An approximately equal quantity is contributed by sulphur in chemically combined ores, of which pyrites (iron disulphide) is the most important on a world basis. Compared with 46 per cent from elemental sulphur, 47 per cent is estimated to come from sulphur in chemically combined ores. The balance of about 7 per cent comes from so-called 'recovered sulphur' which includes sulphur from materials, notably petroleum, natural gas, and coal, which are not primarily thought of as sources of this element. None the less, if these subsidiary sources could be fully utilized, they would come close to meeting total world requirements. Although this is essentially a paper calculation, it suggests the scope which exists for improvement.

In Great Britain, the only considerable source of sulphur apart from coal, is the mineral anhydrite (calcium sulphate), and the next group of photographs illustrates production from this source, in which there has been recent development. As a raw material for the making of sulphuric acid, the value of anhydrite depends on the existence of a local demand for a roughly equal quantity of cement clinker. The process used consists in heating an accurately proportioned mixture of calcium sulphate with materials containing carbon, silica, alumina, and ferric oxide. The sulphate part of the anhydrite is reduced to sulphur dioxide which is driven off as a gas, and the material remaining is a combination of the aluminoferrate and silicates of calcium. These reactions are carried out in a long rotary kiln, into one end of which is fed a dried, crushed, and proportioned mixture of the raw materials, while the other end of the kiln is fired with

pulverized coal. The output of sulphur dioxide is cleaned and converted into sulphuric acid by the 'contact' process, in which sulphur dioxide is oxidized to trioxide in the presence of a catalyst, and the trioxide is absorbed in already made acid.

Kilns installed at the Billingham works of Imperial Chemical Industries, Ltd, before the war are now producing about 100,000 tons of acid annually from anhydrite mined on the site. This rate of output has been more than quadrupled by three recent developments. An extension to the Billingham plant has raised output there to 175,000 tons; the newly completed plant of Solway Chemicals, Ltd, at Whitehaven has a capacity of 90,000 tons of acid, the anhydrite being mined locally, as at Billingham; and the Widnes plant of the United Sulphuric Acid Corporation, Ltd, formed by a group of users of sulphuric acid, will have a capacity of 148,000 tons of acid. These figures, which total 413,000 tons, may be compared with a total British production of a little over 2,000,000 tons of sulphuric acid from all sources during 1954. Inset 6 and 7 show operations in the mining of anhydrite at Billingham; Inset 8, sections of a rotary kiln during installation at Whitehaven; and Inset 9, a completed rotary kiln at Widnes, together with the plant for feeding the prepared raw material into the kiln. As well as this use in making sulphuric acid, much anhydrite is also used at Billingham in the direct production of ammonium sulphate for use as a nitrogen-containing fertilizer. Finely ground anhydrite (Inset 10) is mixed with ammonium carbonate solution, and ammonium sulphate is obtained by direct reaction. In so far as the sulphate moiety is necessary to the distribution of ammonia-type fertilizers in acceptable form, the result is equivalent to an increased production of sulphuric acid.

On a world-scale, pyrites, as has been stated, makes by far the biggest contribution to sulphur supplies, apart from elemental sulphur. The proportion estimated to come from this source in 1956 is 35 per cent; that estimated for 1953 was 37 per cent. In Great Britain, it has been recent policy to increase the number of pyrites-roasting plants, as of plants using anhydrite. Inset 11 shows pyrites mining in Spain, and Inset 12 a modern pyrites

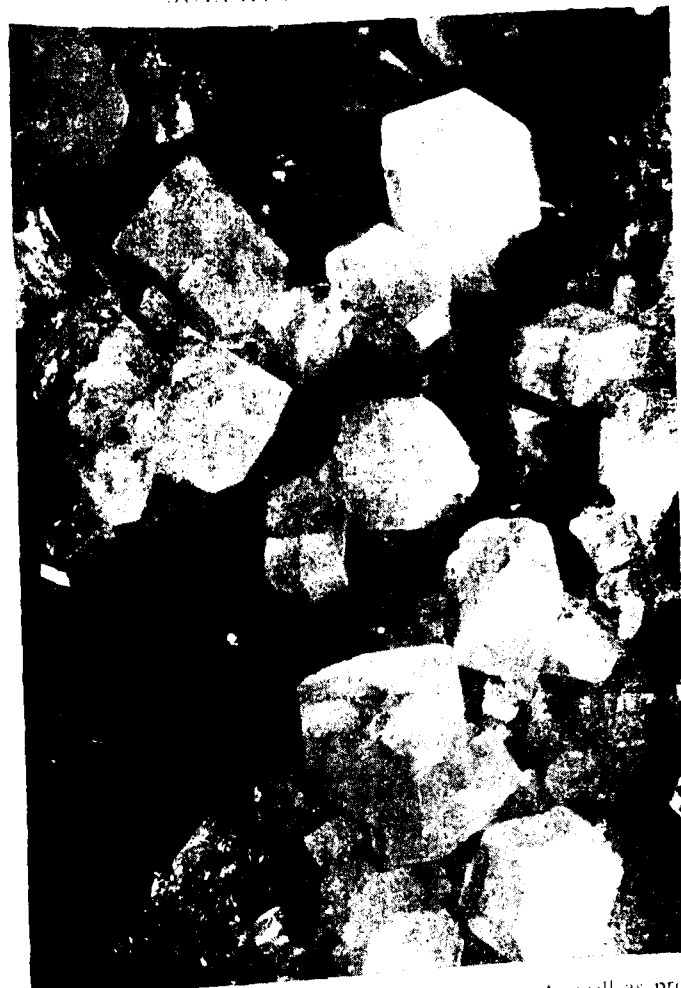
roaster. It seems probable that more than 30 per cent of British sulphuric acid may now be obtained from pyrites. Production from zinc blende (zinc sulphide) is chemically analogous, *but* with the difference that the proportion of sulphur in the ore is lower, and that zinc is the primary product, so that output is limited by zinc requirements.

Most of the remaining pictures are devoted to so-called recovery processes. Both in the case of petroleum products and coal the biggest recovery has come so far from the requirement that gas for public supply must be free from sulphur because of the sulphur dioxide formed in combustion. Thus in the petroleum industry, in 1953, nearly two-thirds of the sulphur recovered came from natural gas, and little more than one-third from petroleum as such. In the United States, the largest sources of natural gas are in western Texas, New Mexico, and Wyoming. Some of these sources are free from hydrogen sulphide, others yield gas containing as much as 42 per cent by volume. Inset 13 shows a recent extension to the sulphur recovery plant at Jumping Pound in Alberta, Canada, where the gas contains about 9 per cent of hydrogen sulphide.

Petroleum offers a potentially bigger source of sulphur, because of the greater quantity used. The sulphur content of crude oil ranges from about 1.5 to 5 per cent according to origin. The difficulty is that the sulphur is chemically combined with the hydrocarbons of the petroleum. Hence it cannot be recovered, if the crude oil is merely separated into fractions by distillation. Chemical processes such as catalytic cracking and hydrogenation lend themselves to sulphur recovery, and because the use of these processes is increasing, so also is sulphur recovery. Inset 14 shows a recovery unit at the Stanlow Chemical Plant. The total annual capacity of British plants for the recovery of sulphur from petroleum is now around 35,000 tons, compared with about 100,000 tons in the United States.

Coal contains a variable amount of sulphur; the average content of British seams now worked is about 1.5 per cent. For the reason already given, the most important channel of recovery is in the production of gas for public supply. In the usual pro-

SULPHUR IN PICTURES



1. Native sulphur: crystals from Girgenti, Sicily. As well as providing the finest museum specimens, Sicily was the great historic source of sulphur. The crystals shown are of about 1-inch width. (By courtesy of the British Museum, Natural History.)



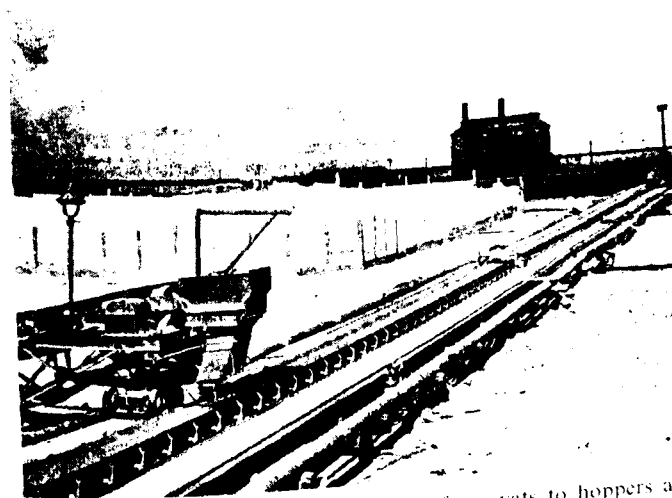
2. Native sulphur: production from Mexican sulphur domes has been added lately to that from the United States. The pipe lines carry superheated water and air under pressure to the wells, and molten sulphur from them.



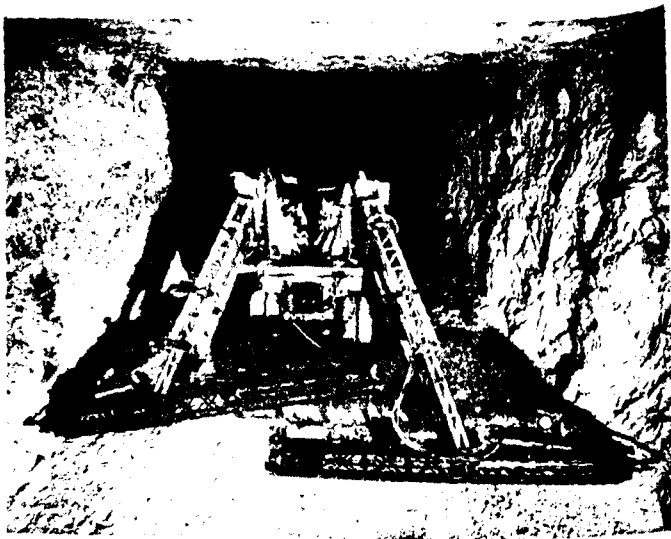
4. Native sulphur: above the Grande Ecaille dome, Louisiana; sulphur barge in foreground. Production from further domes by the Freeport Sulphur Company has lately improved the American sulphur position.



3. Native sulphur: molten sulphur from a well is being discharged into a heated storage tank at a relaying station. (Photographs by courtesy of the Mexican Gulf Sulphur Company.)



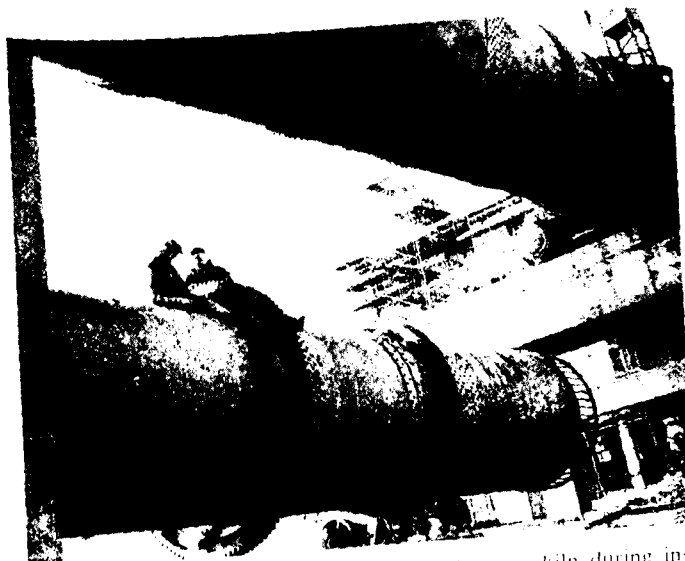
5. Native sulphur: belt carrying sulphur from vats to hoppers at Grande Ecaille, Louisiana. (Photographs by courtesy of the United States Information Service.)



6. Anhydrite mining. Anhydrite (calcium sulphate) is the only important raw material in Britain for making sulphuric acid. This automatic mobile carriage for drilling shot holes is a recent innovation at the Billingham-on-Tees mine.



7. Anhydrite mining: bull-dozer, loader, and diesel truck working together to remove the material released by firing. Output has lately been increased. (Photographs by courtesy of Imperial Chemical Industries, Ltd.)



8. Production from anhydrite: sections of rotary kiln during installation at new sulphuric acid plant at Whitehaven, Cumberland; the anhydrite is mined from local deposits. (By courtesy of Solway Chemicals, Ltd.)



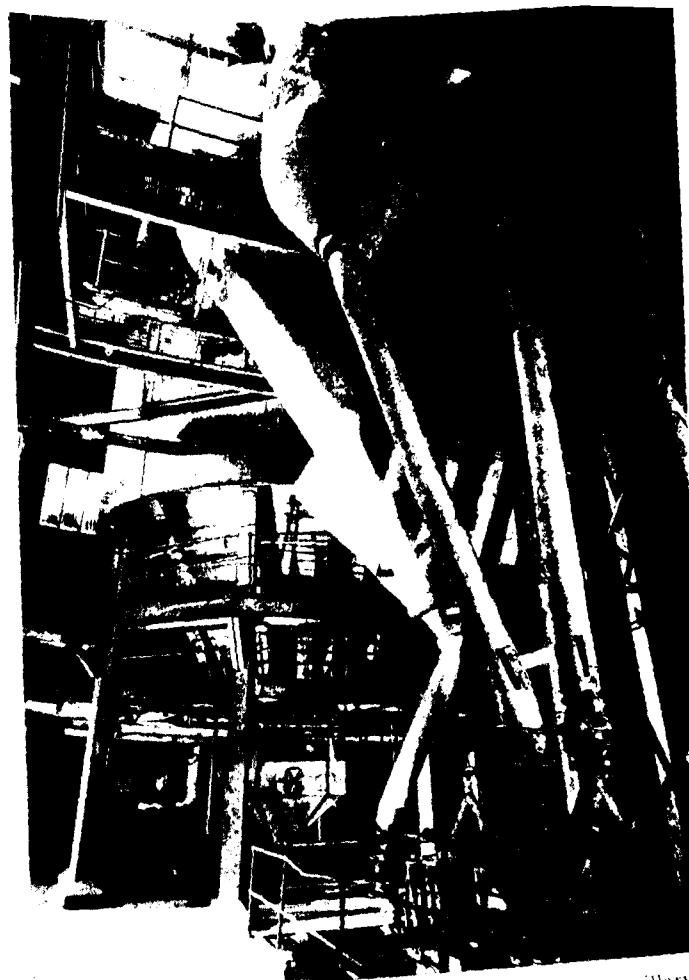
9. Production from anhydrite: new sulphuric acid plant at Widnes, Lancashire. Shale, sand, coke, and anhydrite are fed in at the left and heated in the rotary kiln, sulphur dioxide being driven off. (By courtesy of the United Sulphuric Acid Corporation, Ltd.)



10. Production from anhydrite: at Billingham, Co. Durham, anhydrite is also converted directly into ammonium sulphate for use as a nitrogen-containing fertilizer. The photograph shows the grinding of the mineral in rotary ball mills. (By courtesy of Imperial Chemical Industries, Ltd.)



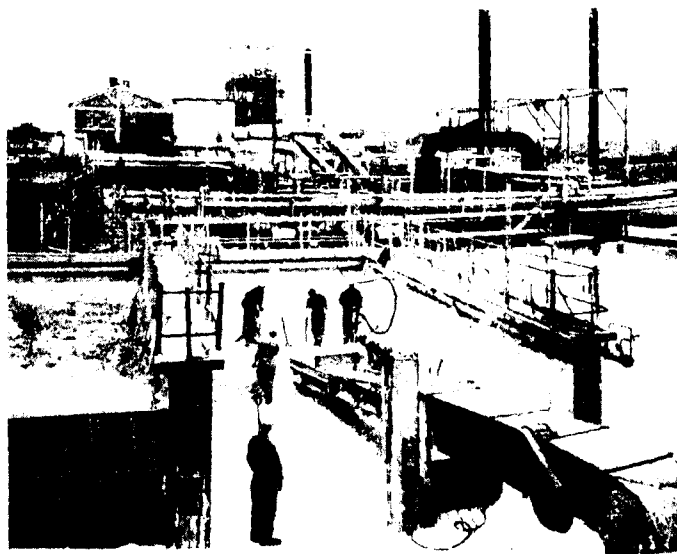
11. Pyrites mining. Iron pyrites (iron disulphide) is the predominant source of sulphur outside the western hemisphere. The picture shows part of the Atalaya opencast workings, Rio Tinto, southern Spain. (By courtesy of Rio Tinto Company, Ltd.)



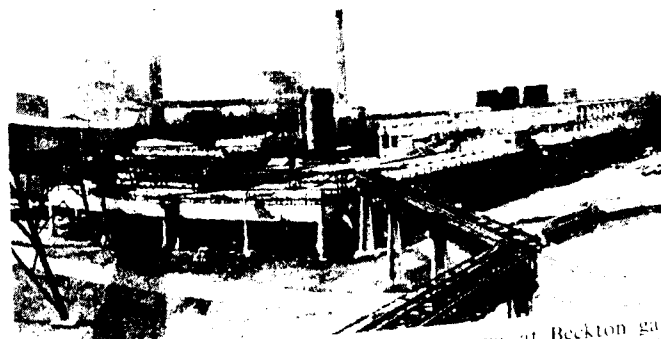
12. Production from pyrites: bottom of pyrites roaster and ancillary equipment in sulphuric acid plant. (By courtesy of Courtaulds, Ltd.)



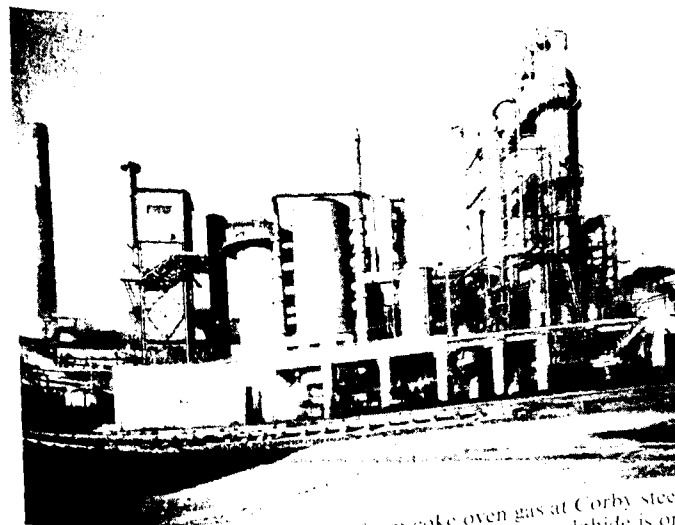
13. Sulphur from natural gas: from a daily output of 60 million cubic feet of natural gas at Jumping Pound, Alberta, 80 tons a day of sulphur are extracted.



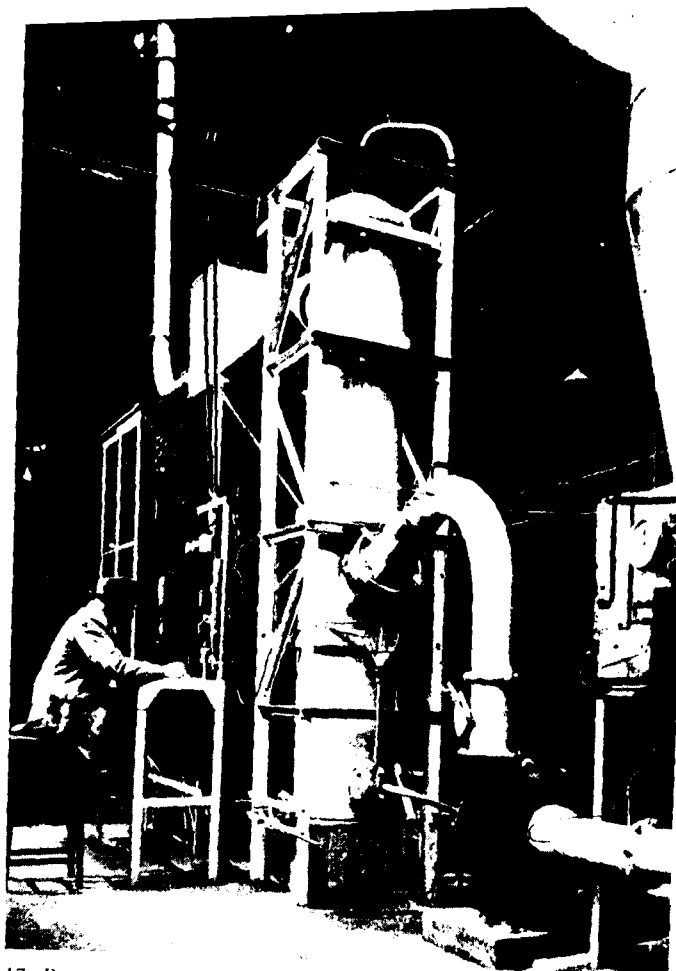
14. Sulphur from petroleum: recovery unit at Stanlow chemical plant, showing how the solid product is handled. The capacity of the plant is about 15,000 tons a year. (Photographs by courtesy of the Shell Petroleum Company, Ltd.)



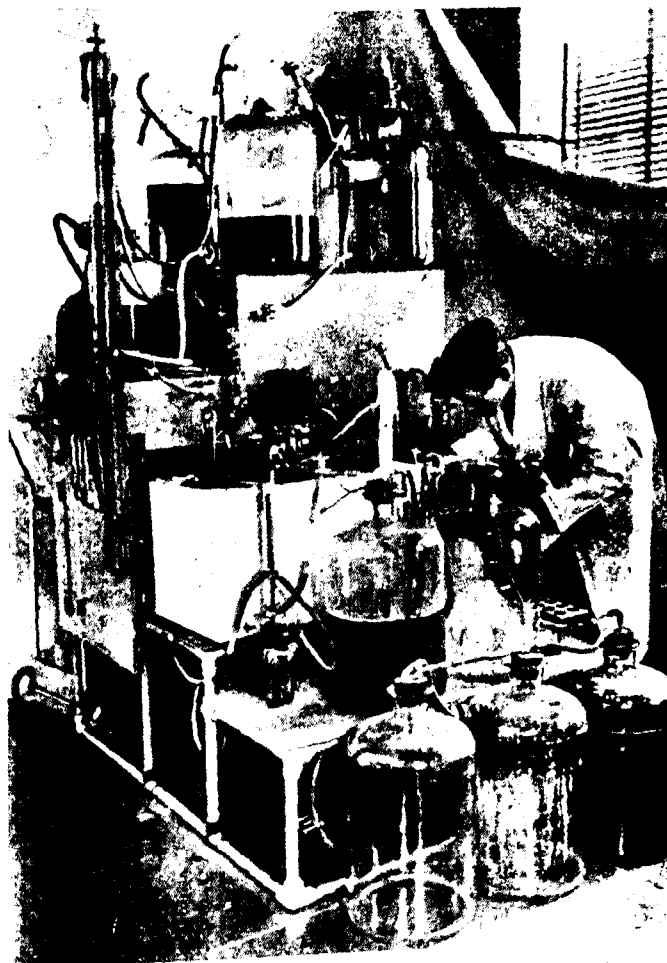
15. Sulphur from coal: recovery from town gas at Beckton gas works. Hydrogen sulphide is absorbed in rectangular boxes ('purifiers'), in which the gas is passed through hydrated ferric oxide supported on wooden grids. (By courtesy of the North Thames Gas Board.)



16. Sulphur from coal: recovery from coke oven gas at Corby steel works. The equipment for the extraction of hydrogen sulphide is on the right, and acid making plant on the left. (By courtesy of Stewarts and Lloyds, Ltd.)

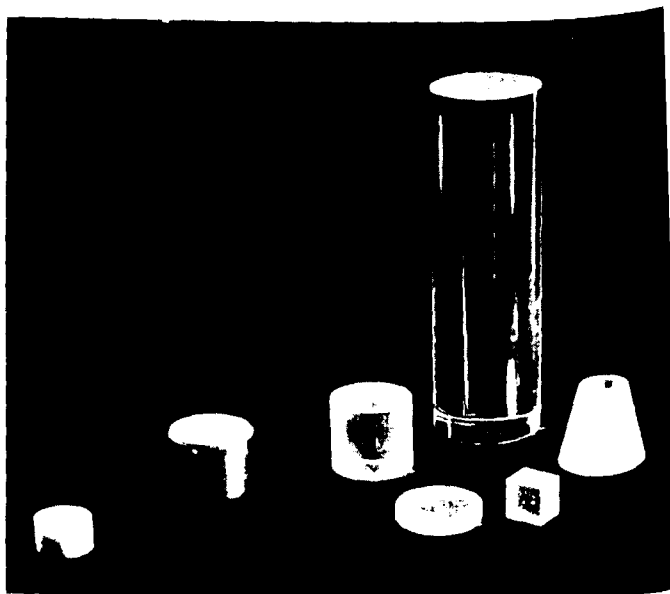


17. Recovery from coal: sulphur dioxide from combustion gases is absorbed by 'ammoniacal liquor' in the Fulham-Simon-Carves pilot plant at the Fuel Research Station, Greenwich. The system is to be given a full-scale trial at the North Wilford power station, Nottingham.

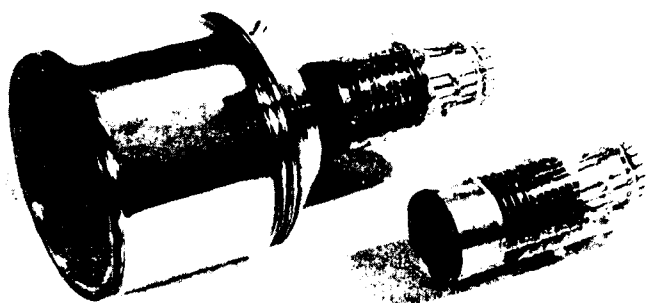


18. Microbiological production from sulphate: laboratory plant at the Chemical Research Station, Teddington, for the continuous culture of sulphate-reducing bacteria, hydrogen sulphide being formed. (The photographs on these two pages are Crown copyright and are reproduced by permission of the Controller, H.M.S.O.)

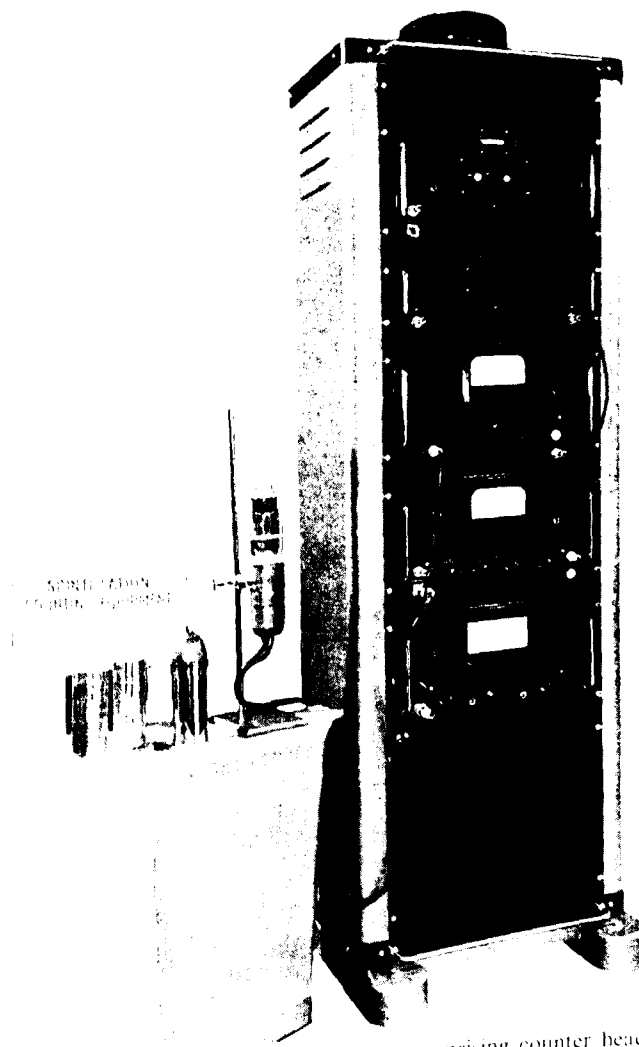
THE SCINTILLATION COUNTER



19. Crystalline and plastic phosphors. (By courtesy of Isotope Developments, Ltd.)

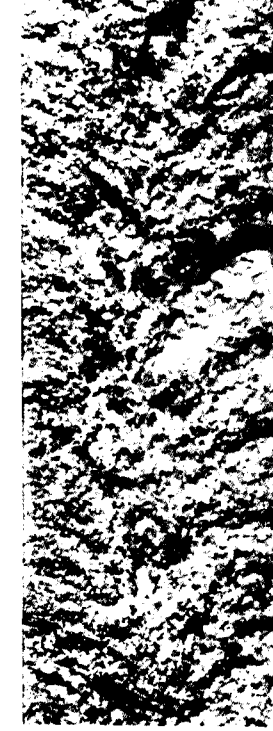


20. E.M.L. photomultiplier type 6097 (right) has a $1\frac{1}{2}$ -in.-diameter photocathode; type 6099 (left) has a larger, 4 $\frac{1}{2}$ -in.-diameter photocathode, and is specially suitable for use with liquid and plastic phosphors. (By courtesy of E.M.L. Electronics, Ltd.)



21. Complete scintillation counting unit, comprising counter head for use with a variety of phosphors and with interchangeable light-tight caps, amplifier and discriminator, 'pre-scaler' and scaler, and regulated power unit. (By courtesy of Isotope Developments, Ltd.)

BRITTLE FRACTURE OF STRUCTURAL STEEL



22. Experimental fractures of ship plate: *above*, 0.339 inches thick, and, *below*, 0.91 inches thick, showing chevron markings. The direction of propagation is from left to right in each case, and the apices of the chevrons point towards the origins of the fractures. The chevrons are coarse and well marked in mat fracture line fractures; (Photographs by Dr C. F. Elam Tipper; the upper picture from The Fracture of Mild Steel Plate by permission of H.M. Stationery Office.)

23. Photograph taken near the starting-point of a service fracture in 1½-in. shell plating (shown on smaller scale); clip attached to plating. The origin of the fracture is at the upper surface of the plate. (By courtesy of the U.S. National Bureau of Standards.)



ARTHROPODS AND TERRESTRIAL LIFE



24. The common wall-spider, *Ciniflo ferox* (male).

PLANT MOVEMENTS



25. Phototropism and geotropism, illustrated in bean plants: *left*, normal; *centre*, bean plants exposed for 24 hours to very weak unilateral illumination from the right, towards which the plants bend (phototropism); *right*, bean plants placed in the horizontal position for 24 hours are bending towards the vertical (geotropism).



26. *Mimosa spegazzinii*: *left*, unstimulated plant; *centre*, same plant some 5 seconds after stimulation (by touch) of the left top leaf, which is now folded; *right*, same plant 1 minute after further stimulation this time by heat applied with a match to the end of the leaf.

cess, the gas is passed through iron oxide (hydrated ferric oxide) supported on wooden grids in long rectangular boxes. The oxide is converted into sulphide; and, if the gas is accompanied by a small proportion of air, sulphur is liberated from the sulphide. Inset 15 shows an array of these boxes at the Beckton gas works; the building behind them contains a further two-tier unit, which enables space to be saved. A further channel of recovery is by the treatment of the gas obtained from coke ovens in the making of metallurgical coke, notably for use in the iron and steel industry. Here the gas is used mainly in the same works, and there is an advantage, but not a requirement, that it be freed from sulphur. Inset 16 shows a plant that has been completed recently at the Corby steelworks. This uses the Collin process in which the hydrogen sulphide is selectively absorbed by quick 'scrubbing' with ammonia liquor, from which the hydrogen sulphide, but not the ammonia, is later driven off. The hydrogen sulphide is used to make sulphuric acid in the plant shown in the left of the picture. The annual output is about 10,000 tons of acid, which represents some 3,000 tons of sulphur. It has been estimated that 67,000 tons of sulphur are wasted annually in Great Britain by burning coke-oven gas without purification, of which upwards of 20,000 tons is contained in gas which is produced on a large enough scale for recovery to be practicable. However, neither in gas works nor in industrial carbonization does recovery of sulphur from the gas mean that the whole of the sulphur has been recovered; the proportion of sulphur in coke is much the same as that in the coal from which it is made, and from 60 to 70 per cent of the weight of the coal remains as coke.

The removal of sulphur from the gases produced in the burning of coal and of solid fuels made from it is thus a potentially bigger source of recovery than anything which is now done in gas works or may be done in the future in the treatment of coke-oven gas. This is a more formidable problem, if only because the volume of gas to be treated is greater, and no general solution seems in immediate sight. In one method, used at the Battersea power station, the combustion gases are washed with river water to which small quantities of chalk are added. With this

method, sulphur is removed – at a price – but is not recovered. Inset 17 illustrates an alternative process, in which the sulphur dioxide is absorbed by ammoniacal liquor and recovered. Although this method is to be given a full-scale trial at Nottingham, it has been pointed out that present supplies of ammoniacal liquor would suffice for the treatment of only about one-sixth of the combustion gases produced by power stations. However, the primary interest of electricity authorities is in generating electricity, and it may be doubted if the approach to the problem has yet been wide enough.

The remaining picture (Inset 18) shows the laboratory plant used at the Chemical Research Laboratory, Teddington, for the continuous culture of sulphate-reducing micro-organisms, using hydrogen gas as the reducing agent. As was reported in *Science News* 34 (p. 119), 'for the process to be economic on a large scale, it would be necessary that a cheap and plentiful supply of hydrogen should be available; or, alternatively, that a continuous process should be worked out for the breaking down into hydrogen and sulphur of the hydrogen sulphide extracted from the culture chamber, so that the hydrogen could be used again'. Since neither condition is immediately satisfied, the process remains rather as an interesting possibility than an immediate contribution to sulphur supplies.

In sum, therefore, it may be said that, while 'easy' (Frasch) sulphur has attractions so long as it lasts, there is both a general and a national advantage in limiting pressure on supplies. The possibilities of rock sulphur are by no means exhausted, supplies of sulphur-containing minerals are plentiful, and the extent of recovery from petroleum and coal is likely to be progressively increased, if not spectacularly.

ACKNOWLEDGEMENTS AND FURTHER READING

Thanks are due to the many organizations and firms which have helped in the assembly of photographs; for information especially to the Intelligence Division, Department of Scientific and Industrial Research, and to the British Sulphur Corporation, Ltd; and for the graph reproduced as Figure 1 to Mr R. D. Cribb.

A Survey of the Sulphur and Sulphuric Acid Position by R. Ash-ton, A. L. Thorogood, and D. Neville-Jones (London, H.M.S.O., 1952) gives a useful survey of the position, primarily in relation to Great Britain, as it appeared at the time of the recent sulphur 'crisis'.

The *Quarterly Bulletin*, formerly issued by the Sulphur Exploration Syndicate, and now continued under the name of the British Sulphur Corporation, Ltd, provides a general source of information on world sulphur, including methods of extraction.

On the problems of sulphur recovery, see:

Special Study of Sulphur Removal and Recovery from Fuels.
London, The Institute of Fuel, 1954.

A. W. H.

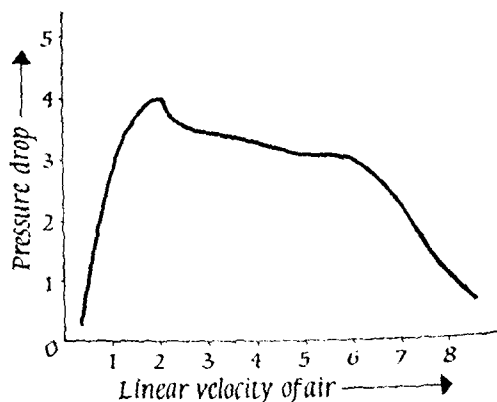


Fig. 2. To show the relationship between air velocity and pressure drop.

Two forms of fluidization can be distinguished. In the aforementioned case the air at first 'bubbles' through the solid bed in a way similar to the bubbling of a gas through a liquid. With increased air velocity the bubbling becomes more violent, and 'streams' and 'ribbons' of the solid are ejected to considerable distances above the bed before returning to it and before true fluidization begins. This is the so-called 'aggregate fluidization'. If, however, instead of air or gas, a liquid, say, water, is used as the fluidizing agent, the oscillation of each solid particle increases in velocity and extent as the bed expands. The porosity of the bed increases accordingly, and finally each particle, unhindered by the others, behaves like an individual. This is called 'particulate fluidization'.

It may now be asked, what practical purpose does all this serve? The general answer to this question is that fluidizing technique has caused a sort of mild revolution in applied chemistry and engineering. In the first place one can imagine the use of this technique for the distribution of all kinds of goods. A column of an adequate size and shape can store and distribute by aeration a great variety of materials, such as grain, flour,

soap powder, cement, etc. Indeed, in the United States, such methods have recently been adopted and carried out on a large scale. Specially constructed railway-cars transport those products. They are fitted with piping and compressed air cylinders and travel as self-contained units. The goods, shipped in bulk, are packed so tightly as to be able to support the weight of a man. None the less, the unloading proceeds so quickly and easily, with only one pound per square inch of air pressure, that the conventional system using bags or containers is gradually falling into disfavour. An American 'Airslide' car is shown in Figure 3.

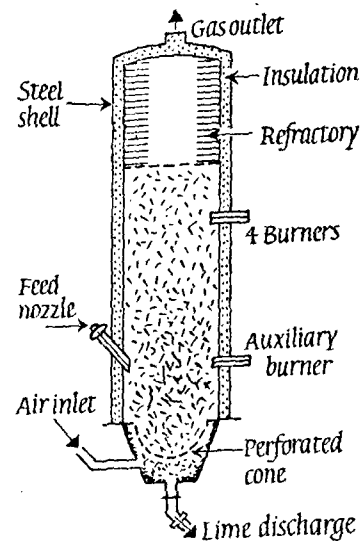


Fig. 3. Fluidizing furnace for the calcination of lime sludge. (Redrawn from Perry's *Chemical Engineers' Handbook*.)

In the chemical industry fluidizing methods have been introduced on a commercial scale into its many branches during and after the last war, and a variety of industrial processes are already carried out in the fluidized-solids reactors. Certain

chemicals are made by reacting gases with a powdered solid, or an intimate mixture of solids, under controlled conditions. This has usually been done by passing the gases over the solids, and, for the mixing of the latter, large, unwieldy, and noisy mixing machines have been used. All these operations can now be performed within a fluidizing column, and this method is far more effective owing to a more thorough interpenetration of the reactants, since the gases have free access to the particles from all sides.

The use of fluidized solids for carrying out high-temperature reactions was developed by the Standard Oil Company in World War II for the catalytic cracking of petroleum 'feedstock' to give a high yield of motor spirit (petrol). Already forty such units have been installed with more to come in the near future. They are vertical cylindrical steel vessels containing finely divided catalysts which are kept in suspension by the flow of vaporized 'feedstock'. The velocity of the vapour varies from 0.5 to 2 feet per second, and this is sufficient to maintain a fluidized bed in the semi-liquid state. The technique is of particular importance when close control of temperature is required, for with this system a better control is possible. In other processes the solid does not play the role of a catalyst, but may serve as heat carrier while gases or vapours are reacting with one another. It has been established that the heat transfer from solid to gas and vice versa in a fluidized column is more efficient than in the conventional apparatus.

A further range of applications depends on the use of 'fluid-solid furnaces'. These are vertical shafts, single, or two or more connected together, fluidized by the ascending gases. They are employed wherever an oxidation can be carried out with air or oxygen at high temperatures, for example, for roasting arsenical, pyritic, and gold ores; sulphides of zinc, copper, and iron; for the calcination of limestone or of calcium carbonate sludge which has been obtained as a by-product. A fluidization furnace used for the latter purpose is shown in Figure 4.

The production of valuable metallic oxides and metals via their halides, for instance, chlorides, is another field where

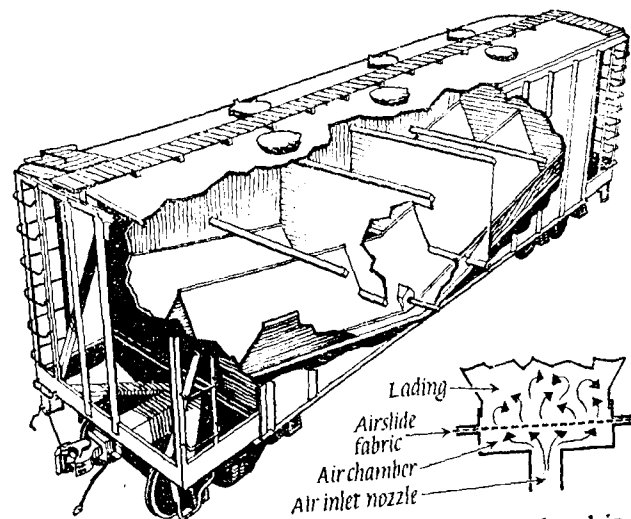


Fig. 4. 'Airslide' distribution by fluidization, as developed in the United States.

fluidization is going to prove of great help. In these cases the ore is treated in a column at high temperature with chlorine gas, which here plays the double role of fluidizing agent and reactant. In a similar way the reduction of metal oxides is feasible; for example, haematite to magnetite, or the conversion of oxides to other compounds, such as copper oxide to copper sulphate.

Quite recently fluidization methods have been introduced into the bituminous and brown coal industries, when a powdered material has to be used for making various valuable gases, such as town gas, or high-power industrial gases, water gas and 'blue gas'. Because of its peculiar properties, the feeding of coal dust is not easy, especially when, as in this case, a quite regular flow is required. On this evenness of flow the whole gasification process depends. There is no form of mechanical device which fully meets requirements, since coal dust lends itself so easily to blockages, 'arching', and similar mechanical disturbances, although electronically vibrated feeders may overcome these

troubles. In the United States it has been found lately that even and regular transport and feeding of the finest coal powder can be successfully brought about by fluidization. For this purpose, the standard apparatus, shown on page 73, has to be altered by attaching to the bottom of the column a vertical, small delivery-tube or even a capillary of, say, 3 millimetres diameter, through which the material is pushed out by the pressure inside the column. It is, however, clear that as the contents of the column diminish, they must be replenished, and that this can be done only after the operation has been interrupted. Therefore, either two separate fluidization systems must be used, and arrangements made for switching quickly from one to the other, or, much better, two columns may be combined in such a way that one serves as a batch feeder and other one as a continuous feeder. They are connected by a capillary at the bottom and form a single operational unit. The first has a hopper for the intermittent supply of coal. From this column the dust passes through the connecting pipe into the second column, and from there into the gasification furnace. For the short replenishing periods, the first column is switched off while there is still enough coal left in the second one to continue the feeding operation. To work efficiently such a feeding system requires a large throughput, that is, the coal dust has to flow through fast enough to prevent occasional stoppages. Astonishingly a capillary of the diameter already mentioned can deliver hundreds of pounds of coal daily. From this brief discussion it can be seen how apt are fluidization methods for the replacement of older and well-known techniques, and it may be expected that many more uses will be found for them.

BRITTLE FRACTURE OF STRUCTURAL STEEL

F. A. FOX

UNQUESTIONABLY the most important and versatile material of engineering construction is mild steel, which is consumed at the rate of millions of tons per annum. There are many reasons why this is so; one of the highest importance is that mild steel is (or should be) not only strong, but also ductile. 'Ductility' implies ability for plastic deformation to occur before fracture. This is important in service, since it represents in effect a reserve of strength to permit a structure to withstand overloads without breaking. If part of a structure is overloaded, it can, if ductile, yield slightly to the loading, and this yielding may, in various ways, have the effect of reducing the overload, so that the structure, even if deformed slightly, may still be not seriously impaired.

'Ductility' is conventionally measured in a tensile test on a steel test-bar by obtaining the amount of plastic extension at fracture and by the reduction of area at the point of failure. A mild steel will usually show a plastic extension of about 25-30 per cent of its original length, and the local thinning down at the point of breakage will amount to perhaps a 60 per cent reduction of area. Ductility is, in essence, the reverse of brittleness. It would become a less important requirement if one could predict with complete certainty every stress which all highly stressed parts of a structure would encounter in every stage of their service; but this is not often possible. One point which is still obscure is the amount of ductility that is required to confer reasonable safety on a given structure. In practice, however, there are clear requirements for ductility as a test of quality of steels and other engineering materials.

Scientific investigation of the deformation and fracture of metals is a matter of considerable theoretical interest, and the newer theories of deformation by propagation of dislocations have given this work a fresh impulse. However, a series of major failures of steel structures, all occurring within the last 20 years, has given a very powerful practical urge to research work on the reasons for the sporadic failure of steels in a non-ductile or brittle manner.

BRITTLE FAILURES

This occasional tendency of steel to brittle failure in service is not in fact new, although it has only recently forced itself on our attention as a phenomenon too expensive to be ignored. Reports appeared in the 70s and 80s of last century referring to the strange failure of steel bars, apparently quite good, when subjected to slight loads. A prelude to the recent failures occurred in March 1938 in Belgium when a bridge over the Albert Canal at Hasselt collapsed into the canal. Eye witnesses heard a sound like a shot and saw a long crack open. All fractures, when later examined, were brittle, some were through welds and others were through solid steel plate away from welds. The bridge was lightly loaded at the time; the weather was cold. Ordinary tensile tests and chemical analysis showed the steel to be apparently quite good both in this case and also in two other bridges of the same type which collapsed in Belgium during the next two years.

One wintry evening in 1943 a welded tanker, the *Schenectady*, lying quietly at her fitting-out pier at Portland, Oregon, having that day passed her acceptance trials, suddenly broke in two with a report which was heard for at least a mile. The deck and sides had fractured amidships and the vessel, held together only by her bottom plating, jack-knifed so that the centre rose out of the water. In the world of ship-builders and steel makers the noise of this failure travelled far and lasted long, but was soon followed by others. Between 1942 and 1952, about 250 welded ships suffered one or more brittle fractures of such severity that they were lost or were in a dangerous condition, and 1,200 more

ships suffered smaller brittle cracks, dangerous but not disabling.

Between 1947 and 1952 there were 15 ships of between 2,000 and 10,000 tons gross which broke in two after major structural failure; of these, four were all-riveted and the rest of welded construction. There have also been brittle failures in structures of other kinds; pressure vessels, pipe-lines, oil storage tanks, and other large outdoor engineering structures have also failed in the same way, the structural steels concerned being perfectly satisfactory according to ordinary specification tests.

These failures have given rise in Great Britain and the United States to one of the biggest research efforts ever devoted to a single problem of materials. These efforts have greatly increased our knowledge of the properties and behaviour of steel, though a single solution to the central problem is not as yet forthcoming, and may not even be possible. The number of variable factors, some well and some less well understood, is quite large; one may perhaps begin with a look at the fundamental principles of deformation as they affect the problem.

TRANSITION TEMPERATURE

It was noted early on that the presence of notches in the steel would have a large effect on the resistance to fracture, and the speed of loading, whether static or impact loading, would also be important. Early work on this problem, and on metals generally, had involved considerable testing with standard notches cut in the samples, the load being applied transversely by impact. It was found that the energy absorbed in fracturing a notched specimen of a given steel was roughly constant at temperatures round atmospheric, but that, as the temperature was lowered, there was a sharp fall in the energy to fracture; the fractures became brittle and were no longer fibrous in appearance, but became bright and 'crystalline'. This temperature at which brittle fracture sets in is called the 'transition temperature'. The onset of brittleness of fracture is not, however, specifically tied to impact loading, since brittle fracture can be induced in a notched steel by slow bending, or slow tension; it can also be induced in un-notched steels. It was clear that the transition temperature is an

important phenomenon; it was noted that, with only slightly different steels, markedly different transition temperatures could be obtained. Some give transition temperatures as low as, say, -70°C , others temperatures of about 0°C , or even higher, and so dangerously near to temperatures which are common in use.

Notched impact test bars are of standard dimensions; if one doubles the width of an impact test bar, one obtains a doubled energy absorption figure – or one should. However, the situation is occasionally met in which a double width bar of apparently good steel gives, on the contrary, a lower absorption of energy than that for a single width bar. Changing other dimensions of impact test bars may also give rise to peculiar results; these facts were obviously of importance, bearing in mind the large size of the steel parts which had failed in service and the small and arbitrary size of laboratory test specimens.

For the crystals of ferrite (essentially iron) of which the bulk of mild steel consists, there exists a definite normal stress required to cleave the crystals across the cleavage planes. This stress appears to be little influenced by temperature and the rate of loading. However, the ferrite crystals can also be deformed by a slip process along certain planes; the slip process is initiated by the reaching of a critical stress, which is increased by decreasing temperature and by increasing rate of loading. If now ferrite crystals are loaded, normal stresses appear on the cleavage planes, and these increase with the load. If the steel is ductile, the specimen can, however, slip and thus transmit load right through the piece before the normal stresses on the cleavage planes have reached a sufficient value to cause cleavage; this process can continue till the possibilities of deformation are exhausted and the specimen cracks. In the brittle steel, however, the ratio between cleavage resistance and slip resistance is lower, and, as the load rises, the normal stresses on the cleavage planes rise until the critical cleavage stress is reached, before the critical shear stress for that steel is attained and the specimen cracks, substantially without slip. Since the normal cleavage stress is unaffected by reducing the temperature while the critical shear stress for initiation of slip rises as temperature falls, it

is obvious that, below some temperature, brittle failure will always occur.

CRACK PROPAGATION

When a crack is once initiated, elastic energy stored in the structure is released, and this energy is used for the propagation of the crack. Thus a brittle fracture may be initiated by some highly local set of conditions prevailing in some small volume of the steel, and propagated by the stress system's stored energy. In the propagated part, there are chevron markings (Inset 22), the apices of which point to the origin of fracture.

If one considers the relative speeds of a stress wave which is equal to the transverse speed of sound in steel, and the high speed of propagation of a brittle fracture, which is probably characteristically about one-third to one-half the speed of sound (speeds of propagation of 2,750 to 6,600 feet per second have been measured), a brittle fracture may continue to propagate for a considerable length even after the external load has been removed.

We can now state that the circumstances in which brittle fracture may be initiated in a loaded part are likely (though not certain) to include (a) low temperature at time of failure, (b) the presence of a notch or notches, and (c) high strain rate. And the signs that brittle failure has occurred will include (a) absence of evidence of deformation at points where the structure has broken, (b) 'crystalline' appearance of the break with chevron markings, and (c) high speed of failure over long lengths of separation.

The impact or shock resistance of mild steel cannot be predicted from tensile strength and ductility; early work to reveal brittleness employed impact loading, and this has led to the opinion, still widespread, that brittle fracture results from impact loading. The failure of the *Schenectady* in quiet water is one disproof of this supposition. Another is given by the evidences of deformation which can be seen in steel plate after it has been pierced by a projectile.

EFFECT OF WELDING

Such considerations have underlined the importance of choice of test methods. Tests of steel, and tests of tests of steel are being made in many places. One factor of importance here is the part played by welding in the initiation and propagation of brittle failure. As we have seen, brittle failure has occurred alike in riveted and in welded structures; however, more failures by far come in the latter than the former class of structure. This general statement is not quite simple, since the total number of riveted and welded structures is unknown, and the number of welded ships built is now about four times as great as ships with riveted hulls and decks. However, it is clear that a large welded structure may be more liable to catastrophic damage than a similar large riveted one, since the former is more 'monolithic', and the high-speed brittle crack is therefore liable to be able to travel much farther and do more damage. Modified practice in the construction of welded ships requires the inclusion of several riveted longitudinal 'crack arresters', similar to butt straps used in riveted construction. Welds in the plating end at a slot behind these arresters. Of 77 ships which suffered serious brittle failures during 1942-52 and which were fitted with arresters, 25 failures did not involve the arresters, 46 involved failures which were stopped by the arresters, and 6 had cracks which restarted on the other side of the arresters.

This question of the effect of welding is again bound up with the question of the detailed mechanism of the effect of notches in producing stress concentrations and so producing cracks under impact or static loading. If one considers only mild steel of a carbon content well below 0.3 per cent, the heat effects of welding on the plate itself will have little direct bearing on this problem. However, the metal deposited in the weld may quite possibly contain minute discontinuities or may have a crystal structure which is weak and prone to micro-cracking under load. These minute discontinuities may then open up to produce a real 'notch', which will then propagate through the plating. Different steels, all of weldable quality, may give very different

performances in virtue of different resistances to crack propagation from a notch of the relevant size.

Test methods must also take into account the large size of the structural parts which fail, and work is going on to find out the minimum sizes of test-piece which will correlate well with actual performance in service. It seems that dimensions are relatively unimportant, except in the presence of a notch; in the presence of a notch, dimensions determine the conditions under which fractures are initiated and propagated. Here again the mechanism of the part played by a notch is decisive, and the notch may be not merely superficial, the result of some surface machining, chisel or file mark, some over-sharp change of section, but may be provided by internal discontinuities such as slag inclusions within the metal; the relative importance of superficial and internal notches will depend not only on their size and shape but on the stress distribution through the sections concerned.

In practice, the idealized standard notch of the laboratory test assumes a wide variety of forms and be due to a variety of causes. Punching and shearing operations may produce fine edge cracks; marks and defects at edges of plates shaped by the flame-cutting process – all these may act as notches of critical effect in high-stress, low-temperature service. One ship is known to have broken in two as the result of failure initiated from so slight a discontinuity as a crater left at the point where a welding arc was struck. Considerations of this sort underline the great importance of good workmanship in the avoidance of brittle failure; equally, good design, in so far as it is concerned with avoiding the presence of sharp corners and sudden stress changes in the structure, is critical too.

RESIDUAL STRESSES

Now the presence or absence of notches is of no consequence at all if the steel is unstressed; the state of stress is therefore a prime consideration. A factor in discussions of stressing which has assumed major importance in this problem is that of residual stresses. These are the internal stresses in the structure arising from processes of manufacture and assembly, and must be

sustained in addition to external loading. Residual stresses are of various types; for example, the material itself, the steel plate, may, as delivered from the steel works, carry internal stresses due to the rolling operation by which it was produced. Such stresses would be due to non-uniformity of deformation through the piece during rolling and non-uniformity of cooling from rolling temperature. Then, the welding operation puts residual stresses into a piece of plate because of differential heating and cooling which produce differential expansion and contraction; the weld area is compressed on heating (the metal tends to expand, but is hindered from doing so by the surrounding cold metal) and is stretched on cooling (it tends to contract, but is again hindered by cold metal). The exact stress system set up depends on a number of factors, including the exact sequence of welding operations, speed of welding, etc.; the vicinity of a butt weld, for example, exhibits compression stresses parallel to the weld and also across it. Where the weld is made under conditions of heavy restraint, the stress system may be complex. Then, again, other assembly operations, such as riveting, may involve residual stressing of a different kind, where slight mal-alignment or inaccurate centring will lead to preloading of parts in order to permit the riveting operation to be performed. The term 'locked-in' stresses is sometimes used to denote the sum of all stresses of this type. In such an internal stress system, a local stress in one sense must, it is true, be exactly balanced by another in the opposite sense; however, such gross 'neutralization' of internal stress is of little importance, since, as we have seen, failure can be initiated in a highly local volume where breakdown occurs.

The magnitude of residual stresses due to welding operations on mild steel plate held in restraint is considerable and may be of the order of, say, 15 tons per square inch or more. Such high stresses have naturally been regarded with great suspicion, but it is by no means certain that they are decisive, even though they may be important in the problem. One might expect that the loading and unloading of a structure, often repeated, would have the effect of reducing the residual stress by causing local

cancellation of positive and negative stresses. However, it has been found that the magnitude of locked-in stresses is not much reduced by the 'working' of a ship at sea. It has been argued from this that, since all welded ships contain locked-in stresses, and since these are not reduced in service, and since only a fraction of the welded ships fail, locked-in stresses are not, in themselves, the main cause of ship failure. However, the last word has not been said on the effect of locked-in stresses.

METALLURGICAL VARIABLES

The problem has been closely examined from the point of view of the variables in steel manufacture which affect the critical transition temperature. The problem would vanish if we could guarantee that each and every batch of structural steel had a transition temperature as low as, say, -65°C . It has been found that for steels, otherwise similar in composition, one which has been fully deoxidized during manufacture will have a lower transition temperature than one which has not been deoxidized. Again, decreasing the crystal size of the iron-rich part of the structure as seen under the microscope (ferrite), heat treating (normalizing) the plate, and the use of a lower final temperature of hot rolling all help to lower the transition temperature. If one increases the carbon or phosphorus contents, the transition temperature markedly rises and brittle failure becomes more a possibility; an increase of manganese has the opposite effect. Ordinary mild steel, as rolled, may have a transition temperature dangerously high at $+15^{\circ}\text{C}$; the same steel, heat treated (normalized) may show a transition temperature at -15°C : low-carbon, high-manganese, deoxidized steel, normalized, will have a transition temperature of about -65°C . Thick plates (say 2 inches) are found to have considerably higher transition temperature than thin plates (say $\frac{1}{4}$ inch) of the same steel. This adverse effect of thick plates, and the beneficial effects of a lower temperature of finish rolling and of heat treatment (normalizing) are, however, only other expressions of the general statement that ferrite crystal size should be small for low transition temperatures, since large plate thickness coarsens,

and normalizing and low temperature rolling refine, ferrite crystal size.

Empirically, we now know therefore how to produce the low transition temperatures we want. We can increase the manganese/carbon ratio, we can normalize, we can deoxidize the steel. These measures can be applied singly or in combination to reduce the liability to brittle failure to a very low level. However, economic considerations in the manufacture and testing of structural steels are naturally very important; changes of composition, of practice, introduction of new procedures or routine tests are all matters of expense which may become very large in the aggregate. The processing of the steel must therefore be made to fit the economics as well as the technicalities of the circumstances; the steel must be made safe enough: it does not need to be made over safe, if that entails extra cost. For steel of a given notch ductility, there will be a given probability of brittle failure, and this will decrease as the notch ductility rises. For British steels, it has seemed that the most ready and effective solution was to concentrate attention on a manganese/carbon ratio which was sufficiently high: a value of about 3:1 is attainable without great difficulty for British steels, and this appears to provide the safety required, at least for steel for thinner plate. For thicker plates – say, 1 inch or more – the other remedies of heat treatment (normalizing) and deoxidation may also be required to restore the necessary safety margin.

The necessity to counter the adverse effects of increasing plate thickness is becoming more important, since, for many purposes, we are today tending to use thicker and thicker plate, as structures become larger and more highly stressed.

It is quite clear that we must understand a great deal more about the mechanics of metal failure generally before we can thread our way with certainty through the mazes of this problem of brittle failure. Work on this topic is being done in many laboratories all over the world. Meanwhile, we can congratulate ourselves in having at least some empirical aids to safety, which are sufficiently reliable to permit us to continue in the structural engineering business.

PLANT MOVEMENTS

A. M. MAYER

MOVEMENT is not usually associated with plants, and least of all with trees or shrubs. If we think of movement in the narrow sense of an entire organism translocating, then we find very few instances in the plant world. Those which we do find are termed uniaxial movements. Perhaps the best known cases are when a unicellular alga such as *Chlamydomonas* moves towards the light or when spermatozooids move towards the egg by means of flagella. Flagella are locomotive organs, whip-like in shape, which the little plant moves to and fro by a process of contraction and expansion, which recalls, but is quite different from, the muscular contractions of animals.

If on the other hand we think of movement in the sense of part of an organism moving, then the examples in the plant world are numerous. Parts of plants, their roots, their leaves, and their stems, all are capable of movements and do in fact move continuously. These movements can be classified in various ways. The classifications which the botanist employs are largely a matter of convenience, by which certain groups of plant behaviour can be related to each other.

Thus it is often customary to refer to autonomic movements. These are movements which are apparently independent of the surroundings of plants. A rough analogy in man would be the continuous beating of the heart, which is quite independent of the environment, as opposed to the response of the hand, which withdraws when it is burned. In the latter case an external stimulus caused the movement. Perhaps the best known example of autonomic movement is the motion described by the tip of the stem of many plants. In plants such as peas or beans it will be seen by visual observation that the free tip regularly describes a circular movement. This circular movement, which is quite

independent of external conditions, is always in the same direction for the same plant species. Thus hops always move with the sun, whereas bindweed invariably moves in the opposite direction. This type of movement is referred to as nutation. Similar types of motion are shown by the tendrils of many plants. In other cases the leaflets of a plant, such as for example *Oxalis*, show a rhythmic opening and closing. It is easy to show the existence of nutation. It is even easy to show that its immediate cause is that the growth on different sides of the plant stem differs. First one side grows fast, then the other, and this goes on in a continuous regular alternation. But this does not help us very much. It is one thing to describe and identify the mechanics of the movement and another to track down and define the operative lever and its mechanism. The situation is analogous to the knowledge we could obtain of a car engine if we did not know that it was fired at regular intervals by a timing mechanism. We would be able to recognize a regular, repetitive movement of the pistons without any apparent cause. It is the nature of the timing mechanism in plant autonomous movements which is still quite unknown. We do not know why first one side grows fast and then the other.

The second type of movement which we meet is, as already mentioned, that in response to changes in the environmental conditions of the plant. Anyone who has ever grown plants, or even looked at them in more than the most cursory manner, must have noted these movements. The bending of plants, grown near a window, towards the light, the downward growth of roots, the erection of the stem of plants bent or partly broken by the wind, all these exemplify the response to changes in the surroundings. Those just mentioned are referred to as tropisms, because there is a directional relationship between the response and its causative effect. Thus the plant organ which is partly in the dark moves towards the light. They differ from movements, described as nastic, in which the response of the plant is independent of the direction of stimulus. For example, flowers of the crocus open when the temperature rises and close when it falls, this being termed thermo-nasty. It must again be stressed that

these distinctions of terminology are based on convenience rather than principle. In all the cases recalled, the actual mechanism of movement is the same: changes in the growth rate of certain parts of the moving organ. While there are many examples of movements of various types, it is more interesting to consider more closely a few that are better known.

As already pointed out, phototropism, or movement towards the light (Inset 25), is widespread. At the beginning of the nineteenth century scientists were already wondering what might be the cause of this phenomenon. Subsequent research has elucidated much about its nature. It has been shown to be a very intricate and fascinating mechanism. Very early on, it emerged that the reaction to light was quantitative. In the same way that the nerves of an animal will respond to different strengths of stimulation, so the plant responds to various amounts of light. At once the question arises, what constitutes different amounts of light. It was found that a certain minimum of light was required to cause a response. This minimum of light must be given during a certain period of time. Thus if we want to cause a plant to bend by illuminating it from one side then we can do this by giving it a certain amount of light continuously for a limited period. We can, however, also give the same amount of light interspersed with dark periods, provided that the dark intervals are not too long. The effect of light is thus cumulative; if the amount is above the minimum, and the intervals between doses are not too long, there is a strict relationship between the time of illumination and the light intensity. Let us assume that illumination for 1,000 seconds with one foot-candle will cause a definite known response in the plant. (A foot-candle is a measure of light intensity.) This response will be a bending of the stem by a measurable number of degrees. If now we illuminate the same plant for 100 seconds with 10 foot-candles or for 10 seconds with 100 foot-candles, the response will in every case be the same. We cannot, however, reduce the time period indefinitely; there is a minimum time during which the impulse must be given to the plant for it to be 'perceived'. Both the minimum time and the minimum of light needed to cause response

are very small. The time may be as little as a fraction of a second and the amount of light required, if given for long enough, not much more than the light from an ordinary torch, in the region of one foot-candle.

The favourite material for the study of phototropism is the coleoptile of the oat seedling. The coleoptile is the sheath enveloping the first leaf of the seedling. It is the coleoptile which first emerges when oat germinates. So much work has been done on it because it is not only especially sensitive to light but also easily worked with and obtained. The light response is complex when high light intensities are reached. Under these conditions there no longer is a straightforward relationship between response and intensity of stimulation. Thus, if we illuminate an oat coleoptile for one minute with 1,000 foot-candles, we do not get 10 times the response given by 100 foot-candles for one minute, nor again does the latter give 10 times as much as 10 foot-candles for one minute. In fact the response, in the sense of extent of curvature, becomes progressively smaller as the light intensity rises till saturation is reached. There is then no further increase in response from increasing the stimulus. Indeed, if the light intensity is made very high, the plant will move away from the light and not towards it. As if all these factors were not enough, the picture is further complicated by the observation that not all light is equally effective in causing phototropism. While red light is almost ineffective, blue light is very effective in causing movement, the response to other colours being intermediate.

Further study of the phototropic mechanism brought out an interesting point. By no means every part of the plant is sensitive to light. Taking again the oat coleoptile as an example, only the top few millimetres were found to be sensitive. If these were covered, the seedling no longer bent in the light. However, it is not this top region which actually bends. The bending occurs further down about one to two centimetres from the tip, showing a clear separation between the region of perception and the zone of response. The next step in the study of phototropism again was fairly easy. It was to show that the actual mechanics

of the bending were due to a difference in growth rate between the illuminated and unilluminated sides of the plant. It was found that the illuminated side grew more slowly than the side removed from the light, and hence curvature must occur.

It was not till the early 1920s that the next real advance took place. Some extremely elegant research work by the Dutch scientist F. W. Went showed that the phototropic response was caused by a chemical substance which was termed hetero-auxin. He showed that by applying this chemical to one side of an oat coleoptile he could imitate its response to light in complete darkness. In fact the difference in growth rate was due to unequal amounts of hetero-auxin on the two sides of the plant. Chemists soon found that hetero-auxin was indole-acetic acid. It was shown later to be produced in the coleoptile tip and to diffuse from there to the zone of response further down, where it was unequally distributed, being present in greater amounts on the dark side. Hetero-auxin is of very widespread importance in plant life, and is not only concerned in the light response.* It has become evident during the last decade that one particular stage of the growth of the plant cell was concerned. When plants grow, their cells first divide and then grow in size. This growth in size is chiefly due to an elongation of the cell, during which it absorbs a great deal of water. It is this stage of cell elongation and water absorption which is effected by hetero-auxin.

So far as phototropism is concerned, two questions remained to be solved. The first is, how does auxin affect growth, and the second, why does light cause it to be unequally distributed on the two sides of the plant. The first question is one on which an increasing amount of research in plant physiology and biochemistry is centred today, as yet without providing a final solution. As regards the second question, however, great advances have been made during the last few years. It seemed pretty obvious that light causes the destruction of hetero-auxin, on the illuminated side of the plant. If this is so, then in the tip of the plant coleoptile, or for that matter in any sensitive region which

*See, for example, *Science News* 18, 32 (1950), and *Science News* 19, 79 (1951).

is illuminated, the auxin appears in different amounts, on opposite sides of the tissue. If this is accepted, then the rest is more or less clear. The hetero-auxin, or plant hormone, diffuses downwards and reaches the zone of response in unequal amounts. Here it causes different rates of cell elongation, which in turn cause curvature. It remained to be shown that light does in fact destroy auxin. An ordinary solution of hetero-auxin is perfectly stable in the light. The effect of light cannot, therefore, be a direct one. An intermediary, which assists in the process of destruction by light, or photolysis, is therefore required. Such substances which are termed photosensitizers are of frequent occurrence. A few years ago the one involved in the photolysis of hetero-auxin was shown to be riboflavin, a compound required by many animals, being a component of the vitamin B complex. Riboflavin had long been known to occur in plants, but here for the first time a definite function for it could be shown. It now became possible to produce an artificial light-sensitive organ. The leaf blades of nasturtium seedlings are sensitive to light, and cause a curvature in the leaf stalk when the blade is unilaterally illuminated. When the leaf blade is removed, no curvature occurs. From young nasturtium plants having only two leaves, the leaf blade was removed. In its place a small glass capillary, containing a mixture of hetero-auxin and riboflavin, was placed on the leaf stalk or petiole. When this capillary was now illuminated from one side, the petiole bent over, in the same way as it would do if the leaf were illuminated. Thus for the first time an almost complete picture of phototropic curvature was obtained.

Very recently an additional discovery has been made, whose significance has yet to be fully evaluated. Some of the yellow pigments of plants, the carotenoids, interfere with the phototropic response. Thus if we add to the mixture in the capillary placed on the nasturtium seedling a suitable amount of carotene, the leaf stalk will no longer bend. The carotenoids can apparently prevent the destruction by light of auxin. It seems that they compete with the riboflavin for the light. If they absorb sufficient light, the riboflavin can no longer mediate the destruction of auxin. It is still too early to assess the importance of this

discovery from the physiological point of view. It is possible that the carotene is, in fact, responsible for the unequal transmission of light through the oat coleoptile. Without carotene the light would reach both sides of the coleoptile in equal amounts, as this sheath is only about two millimetres thick. The carotene, however, may provide a filter, which causes a suitable drop in light intensity across the coleoptile. This view still requires additional experimentation before it can be finally accepted.

A second type of plant movement is that referred to as geotropism. Plants, like every other object on our planet, are under the influence of gravity, and the various parts of the plant have developed specific responses to this stimulus. Thus roots, generally speaking, grow downwards, towards gravity, and are hence referred to as positively geotropic. In the same way, stems which grow upwards, away from gravity, are negatively geotropic. These responses are of universal occurrence. When an entire plant is placed horizontally, its stem will quickly turn upwards (Inset 25) and its roots downwards. This can be easily shown, but it is less easy to prove that gravity is the operative stimulus. Obviously we cannot remove gravity as we can light. We can only do the next best thing, that is, expose all sides of the plant equally to gravity. This is done by placing a plant horizontally on a wheel which slowly revolves in the vertical plane around a horizontal axis. In this way the plant is slowly turned so that each side in turn is exposed to gravity to the same extent. When this is done, the plant no longer shows any upward or downward curvature. It just grows straight along.

Many of the observations which have been described for phototropism are also true for geotropism. Thus the movement is due to unequal growth, which is caused by an unequal distribution of auxin. The stimulus must be of a certain minimum length and intensity in order to cause a response. The exact perception of the gravitational stimulus is, however, less clear. Many plants contain in the root tip large starch grains which move, under the effect of gravity, within the plant cell. It has been suggested that these starch grains, called statoliths, are the receptors of the gravitational impulse. Although the position of

the statoliths frequently parallels the final position which the root assumes, this is by no means always so. Moreover even plants which have no statoliths respond to gravitational stimuli. But whatever the immediate receptor of gravity may be, it is far from clear what is the next step in the reaction. Why should gravity, however perceived, cause eventually an unequal distribution of auxin? It is worth pointing out here that the reaction of roots to auxin differs from that of stems. While in the stem the side containing most auxin grows fastest, the opposite is the case in the root. This explains why, when a plant is placed horizontally, the stem turns up and the root down. The auxin in this case concentrates on the lower side of the plant. This causes more rapid growth of the cells of the lower side of stem tissue, causing them to bend upwards, and faster growth on the upper side of root tissue, causing it to bend downwards. Hence the overall geotropic response is obtained. It has been suggested that the reason for the unequal distribution of auxin in the horizontally placed plant is the creation of a potential difference across the plant. The under side becomes positive and the total potential may build up to some 10 millivolts. It is claimed that this causes an accumulation of auxin on the lower side of the plant because of its polar flow towards the positive pole. This explanation of the redistribution of auxin is at best an over-simplification of the situation. There are many aspects which it leaves entirely unexplained. The precise reason for the redistribution of auxin in response to a gravitational stimulus is not known.

Perhaps the most striking of all plant movements is that of the sensitive plant, *Mimosa pudica*. This plant is sensitive to touch, impact, electrical impulses, and heat. While geotropic and phototropic responses are generally slow, taking many minutes, the response of the sensitive plant to any form of shock is almost instantaneous. *Mimosa* has compound leaves made up of leaflets, which in turn are made up of pinnules. When the plant is slightly stimulated, by touch, the pinnules begin almost instantly, within seconds, to fold up. When the stimulation is somewhat more vigorous, this is followed by a drooping of the whole leaflets at the junction to the leafstalk, or petiole. When a really

vigorous stimulus is provided, for example, by applying a match to the end of one pinnule, the petiole itself droops at its junction to the stem, and this is followed by the closing and drooping of neighbouring leaves, till all the leaves of an entire branch have folded up. This whole process is extremely fast, and within a minute all the leaves of the branch will have closed. The sequence of closing is important. In the stimulated leaf, as already stated, first the pinnules, then the leaflets, and then the leaf stalk or petiole close up. In the neighbouring leaves, however, the process is reversed. Here first the petiole drops, followed by the leaflets, and only as the last step do the pinnules fold up in pairs. (Inset 26.)

The movements which have been previously described were caused by differential growth. The movement of the sensitive plant, which is frequently termed thigmonastism, is different in mechanism. The pinnules are connected to the leaf rachis, and this is connected to the petiole by hinges which are called pulvini. In the same way the petiole is attached to the stem by a pulvinus. These pulvini contain cells which, in the normal state, are completely filled with water, that is, they are turgescant. They can, however, lose this water to neighbouring cells and intercellular spaces. These cells are chiefly situated on the lower half of the hinge. When, therefore, a sudden loss of water occurs, the lower half of the hinge loses its rigidity and can no longer support the weight of the leaflet or leaf attached to it, which as a result droops downwards. This is the operative mechanism by which the parts of the mimosa plant move.

In view of the extreme rapidity of the response of this plant, which almost suggests the nervous response in animals, it was very soon suggested that a chemical substance passes through the plant very rapidly. It appears, however, that there is more than one way of transmitting the stimulus. One is very fast and passes in one to two seconds from pinnule to petiole pulvinus. This pathway exists only through living cells and cannot be transmitted through dead tissue. The second is somewhat slower and takes up to a minute or two. It can pass through dead tissue. It is apparently the method of transmission from leaf to leaf

along the stem. Although it seems fairly obvious that a chemical substance is involved, this has still to be shown conclusively. Moreover, even if it is shown, it still remains to explain why such a chemical should cause a sudden loss of turgor of cells in the pulvinus. In fact this most puzzling of plant movements is still as far as ever from having found its final explanation.

There are a multitude of other plant movements in response to a variety of stimuli. Movements evoked by chemicals, temperature, and water are among the better known. In none of these has the nature of the movement been fully established. Despite the large accumulation of factual knowledge, the real mechanism of the various plant movements is one of the least understood aspects of plant physiology. The whole field presents a challenge which plant physiologists are at present taking up again with all vigour.

FURTHER READING

The development of the concepts of plant movements is reviewed in:

Weevers, Th. *Fifty Years of Plant Physiology* (chapters 9, 11, 12, 13 especially). Amsterdam, Scheltema and Holkema, 1949.

More recent advances are summarized in:

Schrank, A. R. 'Plant Tropisms'. *Annual Review of Plant Physiology*, 1, 59 (1950).

Brauner, L. 'Tropisms and Nastic Movements'. *Annual Review of Plant Physiology*, 4, 163 (1954).

ARTHROPODS AND TERRESTRIAL LIFE

J. L. CLOUDSLEY-THOMPSON

It is generally agreed that life must have begun in the sea and may have remained confined to the waters for more than half the period that has elapsed since its origin. Despite the instability of climatic conditions on land in comparison with those in the oceans, the terrestrial environment provides abundant vegetation as a basis for food chains, a copious supply of oxygen and the possibility of rapid movement. Consequently when plants became firmly established on dry land they were probably soon followed by various elements of the fauna. Life on land entails a number of problems for animals: larger forms require structural support, respiratory organs must become modified for air breathing, and there is no longer the surrounding water into which toxic excretory products can freely diffuse; while mechanisms for the conservation of water and the maintenance of a constant internal medium must be evolved.

In this article I shall discuss some of the ways in which a few problems involved in the conquest of the land by the Arthropoda have been solved. That these problems are by no means easily overcome is shown from the fact that several invertebrate phyla are almost entirely marine while others as yet possess few terrestrial species. Indeed, the arthropods have been unusually successful in exploiting the terrestrial habitat, and, by adaptation, have succeeded in establishing themselves in nearly all the habitable corners of the earth. Each aspect of their adaptation to life on land affects and is affected by other aspects. For example, it might appear a fairly simple matter for an organism to eliminate water-loss by the evolution of an integument completely impervious to water-vapour: but such an integument

would also be impervious to oxygen and carbon dioxide. A respiratory mechanism has therefore had to be evolved which permits gaseous exchange without excessive water-loss. If the integument is rigid and provides support, then growth becomes impossible except by moulting, and this limits size. A size limitation is also imposed by respiratory mechanisms that depend upon the rate of gaseous exchange. Nutrition and excretion, too, are closely concerned with water conservation, as are behaviour and ecology.

AVOIDING WATER-LOSS

There are two obvious ways in which small animals can escape becoming desiccated on dry land. One is to avoid dry places and to remain most if not all of the time in a humid environment; the other, to evolve an impervious integument. Both methods have been exploited by the Arthropoda and each has its drawbacks and advantages. Indeed, on the basis of this character the terrestrial members of the phylum can be divided roughly into two main ecological groups, the first of which includes woodlice, centipedes, millipedes, and their allies which lose water rapidly in dry air, and the second the arachnids and insects which are covered with a layer of wax that renders them comparatively independent of moist surroundings. In conditions warmer than they are likely to encounter in their natural surroundings, however, the latter group begin to lose water quickly by evaporation above a critical temperature (Figure 1).

Behaviour mechanisms

Physiological and morphological adaptations alone would be insufficient to support the life of any free-living animal. Orientation and behaviour mechanisms must also be evolved to retain organisms in environments to which they are suited, to enable them to find food, mate, and indeed to carry out the innumerable functions essential for the continued existence both of themselves and of their species. Humidity is, of course, an environmental factor of prime importance to animals lacking a water-proof integument. Woodlice are very sensitive to humidity

gradients and aggregate in areas of high humidity. The mechanism by which this occurs is twofold: firstly, the animals show a decrease in activity and speed in moist air, and, secondly, they turn more frequently in damp places so that once they have arrived in a moist situation they tend to remain there. This re-

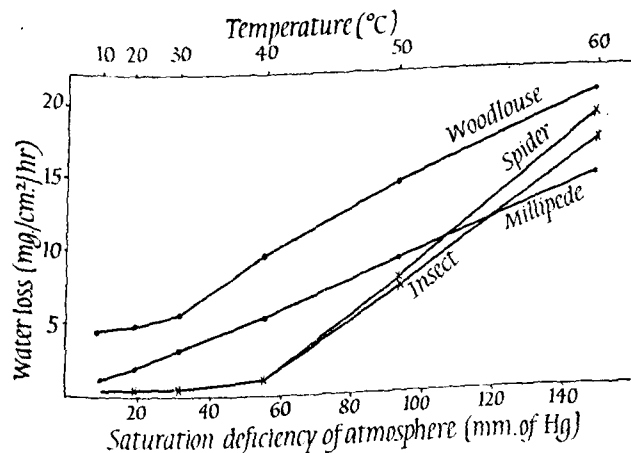


Fig. 1. Rate of water-loss in dry air at different temperatures, and corresponding saturation deficiencies, of a woodlouse (*Porcellio*), millipede (*Oxidus*), spider (*Lycosa*), and insect (*Pieris* larva). In the woodlouse and millipede the rate of water-loss is proportional to the saturation deficiency of the atmosphere, but in the spider and insect it is negligible below about 40° C, the critical temperature at which their epicuticular wax-layers become porous. Rate of water-loss is expressed in milligrammes per square centimetre of surface area per hour.

action to the relative humidity of the air is also combined with avoidance of light and a reaction that causes them to make contact with as much of their surroundings as possible. The result of this is not only that the animals enter narrow crevices, but they bunch together and consequently protect each other from evaporation. However, a reversal of the response to light occurs

when woodlice become somewhat desiccated, so that if the daytime habitat should dry up they are not restrained there; they die, but become attracted to light and are then able to wander in the open until they find some other damp, dark hiding place where they again become photo-negative.*

Somewhat similar responses have recently been shown to occur in myriapods which also lose water rapidly in dry air. Some species may even tend initially to move towards dry air but the reaction is gradually reversed as desiccation proceeds. Reactions to moisture, however, are of far greater importance in the lives of these creatures which seldom come to rest except on a damp surface, than are responses to atmospheric humidity.

The influence of water as distinct from air humidity has not been widely investigated in any arthropods; but it has been shown to be important to wireworm larvae whose wax layer becomes abraded if rubbed by the soil in which they live. In these animals orientation is achieved by two different mechanisms: first, the larvae are more active in dry than in moist air, and secondly they show a directed response to low humidities. Thus they migrate rapidly out of dry soil and aggregate in damp areas.

The reactions to humidity which are shown by insects and arachnids possessing a wax layer usually result in the avoidance of moist air and are therefore not easily correlated with any specific physiological requirements. Examples of such behaviour are afforded by locusts, mosquitoes, lice, cockroaches, houseflies, and mealworm beetles. It is interesting to note, however, that even in these animals the responses may be reversed with an adverse water balance.

Although humidity is the most important factor of the environment to woodlice and myriapods, they are unable to find their way directly to damp places. Instead they are merely repelled by drought. Nevertheless this stereotyped and curiously negative behaviour is surprisingly effective in preventing them from wandering away from their normal habitats: but it does raise the problem of how dispersal can take place and new habitats become colonized.

*See discussion in *Science News* 28, 76 (1953).

There are a number of cases on record of millipedes, sometimes accompanied by centipedes and woodlice, migrating in vast armies. Occasionally they have crossed railways and been squashed in such numbers that locomotives have been impeded and sand has had to be strewn on the lines before their driving wheels would grip. At other times cattle have refused to graze on invaded pastures, wells have been filled with drowned corpses and workmen cultivating the fields have become nauseated and dizzy from the odour of millipedes crushed by their hoes. Such mass migrations, however, are of rare occurrence and local in extent, so that their net effect on distribution is probably negligible. An explanation of the problem has been suggested as a result of recent work which has shown that the restraining mechanisms are somewhat relaxed during the night when the temperature falls and the humidity of the air increases. Thus it is at night that these creatures are enabled to disperse themselves and overcome the restrictions inherent in the physiology of their integuments.

The significance of the absence of a cuticular wax layer and the development of strongly marked reactions to environmental humidity gradients can be explained in terms of the necessity for adequate oxygen supply in the absence of suitable respiratory organs. An alternative suggestion has recently been proposed, however, in the case of woodlice, according to which, although different species can withstand terrestrial conditions of drought to varying degrees, probably all spend most of their lives in an atmosphere saturated with water vapour and differ merely in the length of time that they are capable of surviving in dry air. The conquest of the land by the ancestors of woodlice probably took place across the littoral zone, for shore-dwelling species are still the most primitive in structure. Now animals crossing this zone may well be subjected to high temperatures, and the ability to lose heat by evaporation of water may have had considerable survival value.

It would be mistaken to regard the absence of a cuticular wax layer as a primitive characteristic, however, although the forms that lack one are so restricted in their choice of environment that

they cannot be regarded as entirely successful land animals. Rather, it seems that a particular method has been exploited for surviving the conditions of life on land. At the same time, although the control of body temperature by evaporation cannot occur in the arthropods that have the otherwise overwhelming advantage of being comparatively independent of moist surroundings, the temperature of the thorax of flying insects such as locusts may sometimes exceed the critical temperature at which their cuticular wax layer becomes porous, so that local evaporation and consequent cooling may take place.

Impervious integuments and respiration

The integument of all arthropods seems to be basically similar although it may vary greatly in complexity in different groups. An exceedingly thin outer epicuticular layer of a substance known as 'cuticulin' is always present, but only in insects and arachnids does this support the impervious layer of wax to which their success on land is largely due. The insect cuticle (Figure 2) is light in weight, rigid, impervious, and capable of endless morphological and physiological modifications. Its inner layers, known as exo- and endo-cuticle, differ from the epicuticle in that they are composed largely of chitin, a nitrogenous polysaccharide having the empirical formula $(C_8H_{13}O_5N)_x$. Chitin itself probably consists of long acetylated glucosamine residues, and is soft and flexible. The cuticle owes its rigidity to a brown or amber substance known as 'sclerotin' which binds together the constituent fibrils or micellae. Sclerotin is formed from proteins which have been hardened and made insoluble by the action of quinones; a process analogous to the tanning of leather. The same kind of reaction occurs in some of the epicuticular layers which, however, lack a chitinous foundation; while cuticulin itself is believed to be a condensed, fatty lipoprotein tanned by quinones.

An alternative method of hardening the cuticle occurs in Crustacea whose endocuticle is heavily calcified and brittle, as in the shells of lobsters and crabs. Calcification, however, adds greatly to the weight of the integument and so is only really

satisfactory in aquatic forms. In millipedes the cuticle is hardened both by sclerotization and by calcification and is extremely strong, an adaptation perhaps correlated with burrowing and pushing.

The process of tanning or sclerotization referred to above was first clearly demonstrated in the egg case or ootheca of the

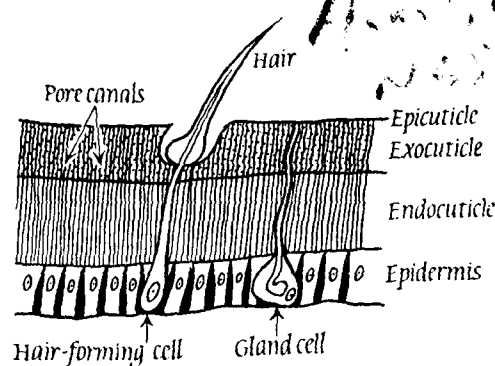
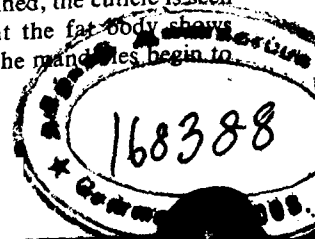


Fig. 2. The insect cuticle, semischematic (after Wigglesworth).

cockroach. Here a watery solution of protein is secreted by the left accessory or colleterial gland, while the right gland secretes a solution of protocatechuic acid, a dihydroxyphenol. On mixing the solutions, the protein becomes hard and brown. This is because polyphenol oxidase, an enzyme present in the secretions, oxidizes the protocatechuic acid to its corresponding quinone and this in turn tans the protein. Adjacent polypeptide chains in the protein molecules become linked through their reactive amino groups which are thus blocked. The same tanning process occurs in the protein impregnating the chitin of the exocuticle as well as in some of the non-chitinous epicuticular layers. If a newly moulted insect is examined, the cuticle is seen to be colourless and transparent so that the fat body shows through. First the hardest parts such as the mandibles begin to



darken, and within an hour or two the whole cuticle becomes reddish-brown and finally blackish.

Various modifications of the integument allow for flexibility and extension. For example, the hardened exocuticle may be broken up into wedge-shaped pieces at intersegmental membranes to permit bending, or may be absent in bugs and other insects that take large meals at irregular intervals (Figure 3). In these, the epicuticle is much folded to allow for the expansion of the abdomen during feeding. In beetles, the hardness of the integument is often enhanced by the presence of a network of strands known as 'balken' embedded in a homogeneous matrix of the endocuticle. Certain ticks and insects are able to take up water vapour from the atmosphere even when in an unsaturated environment without losing their resistance to desiccation at low humidity. The mechanisms by which this takes place, however, are not fully understood.

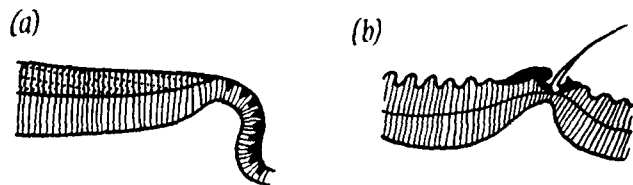


Fig. 3. Examples of flexible cuticle: (a) intersegmental membrane of a beetle, with the exocuticle broken up into wedge-shaped pieces to permit bending; (b) extensible cuticle of a bug, exocuticle being absent except around the bristle, and epicuticle much folded to allow for expansion.

A number of authors have stressed the importance of moisture on the distribution of spiders. The conflict between the incompatible requirements of respiratory exchange and the prevention of water-loss has recently been demonstrated by the writer by a comparison of two common British species *Ciniflo ferox* (Inset 24) and *C. similis*, both of which have a cuticular wax layer with a critical temperature at about 30° C, above which they quickly lose water by evaporation in dry air. At

lower temperatures, however, the rate of water-loss in *C. ferox* is almost double that of *C. similis*. *C. similis* 'tires' very rapidly when forced to run at full speed without stopping and is almost always overcome in fights between evenly matched individuals of the two species. Both species can run for long periods when supplied with oxygen and it is suggested that the greater stamina of *C. ferox* depends upon a proportionately larger respiratory surface acquired at the expense of greater dependence upon environmental humidity.

The primitive respiratory organ of the Metazoa is the skin, but special respiratory organs have been evolved in all but the smallest and simplest of the arthropods. In insects whose integument is impermeable, a special system of minute air tubes or 'tracheae' leads into even finer 'tracheoles' that carry oxygen directly to the tissues where metabolic processes take place. The Arachnida possess both tubular trachea and hollow spaces – lung-books – which communicate with the external air by a small pore and contain a number of respiratory leaves and lamellae (Figure 4). Scorpions have lung-books only, while spiders are passing through a primitive lung-book stage from

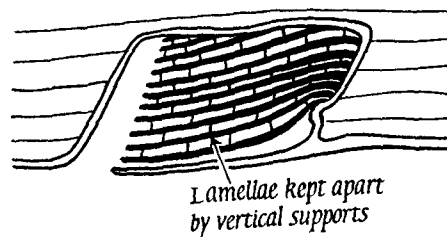


Fig. 4. Lung-book of a spider as seen in section.

which none have yet emerged. Two pairs of lung-books without tracheae occur in the more primitive families while most others have an anterior pair of lung-books and a posterior pair of tracheae. Lung-books provide a localized respiratory area from which oxygen is distributed by the respiratory pigment haemo-

cyanin in the blood. Insects possess no respiratory pigments because all their tissues are supplied with oxygen directly by the tracheoles.

The entrance to both lung-books and tracheae is guarded by spiracles closed by means of special muscles (Figure 5). In this way the minimum quantity of air necessary for respiratory purposes is allowed to circulate and consequently water-loss is much reduced. When an insect or arachnid is placed in an atmosphere containing an abnormally high proportion of carbon dioxide the spiracles are kept open to their fullest extent and the rate of water-loss in dry air is then considerably increased.

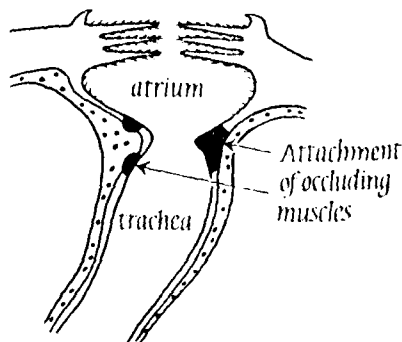


Fig. 5. Spiracle of a butterfly larva as seen in section.

WATER CONSERVATION

Growth, Moulting, and Metamorphosis

In all organisms growth tends to be a cyclical process, periods of rest alternating with activity. H. G. Wells used this idea in his novel, *The Food of the Gods*. In no animals, however, is it more marked than in the arthropods, whose development is punctuated by a series of moults or ecdyses, each of which is preceded by a period of active growth and followed by one of apparent inactivity. All parts of the integument are moulted together, independently of nerve supply. The stimulus for this

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is harmonic in nature, metamorphosis too being controlled by hormones.

During moulting the old endocuticle is digested by enzymes in the moulting fluid and withdrawn in solution through the 'pore canals'. These are numerous fine tubes, sometimes numbering over a million per square centimetre of surface, which penetrate the exo- and endocuticle. A new epicuticle is always laid down beneath the old one before moulting takes place, so that this process is accompanied by little loss of water and a high degree of economy is achieved.

Nutrition and Excretion

Woodlice and millipedes are almost entirely vegetarian, while centipedes and arachnids, with the exception of certain mites, are primarily carnivorous; but different species of insects seem to be able to thrive on almost any kind of organic matter capable of supporting metabolism. In most, water conservation is of prime importance, but blood-sucking insects, those that feed on the sap of plants and some aquatic species, may have to eliminate considerable quantities of water in a short period of time. Thus the bugs that feed on the copious watery juices of plants possess a characteristic filter-chamber that allows fluid to pass directly from the anterior to the posterior regions of the midgut, short-circuiting the main digestive region of the alimentary canal. This permits the absorption of very large quantities of food, necessary on account of the scarcity of nitrogenous material in plant sap, without excessive dilution of the digestive enzymes.

The function of excretion, like that of osmotic regulation, is the maintenance of a constant internal environment within the animal body. An aquatic animal surrounded by a large volume of water can excrete toxic nitrogenous compounds with impunity, but the necessity for water economy in terrestrial forms generally requires the excretion of some compound less poisonous than ammonia. More than half of the soluble non-protein waste products of woodlice are excreted in the form of ammonium hydroxide, but the retention of this primitive character has

been attended by a general suppression of nitrogenous metabolism. In higher insects, as in birds and reptiles, uric acid is the chief excretory compound. This substance has in the first place been evolved in response to the conditions of embryonic life. Uric acid metabolism is correlated with the development of eggs enclosed in a comparatively impermeable shell. Ammonia is too toxic as we have seen and is only suitable as an excretory compound for eggs developing in water: it cannot be converted to urea as this would cause uraemia, and in excess upset the osmotic relationships of the egg. Hence uric acid, a non-toxic, highly insoluble substance, has been evolved which can accumulate within the egg without causing any ill effects.

When a chick hatches, the uric acid produced in embryonic life is left behind in the allantoic membrane. The newly hatched insect larva soon gets rid of the uric acid which has accumulated during its embryonic life, and shortly after emerging from the pupa, drops of liquid, the 'meconium', are discharged from the anus of the adult or imago insects and these contain the waste products of pupal metabolism. The chief excretory compound of the arachnids is guanin which, like uric acid, is very insoluble.

The principal excretory organs of insects are known as Malpighian tubules, named after Malpighi who first noticed them. Uric acid crystals are excreted into the lumen of the tubules and pass into the hind gut, where water is withdrawn by the rectal glands so that a dry pellet of faeces and uric acid containing practically no water is passed to the exterior. Other excretory compounds may be stored and are frequently utilized in the form of pigments. The composition of the excretory compounds, however, depends ultimately upon the food and the amount of water available. For example, blood-sucking insects excrete a clear fluid containing excess water and salts for an hour or so after a meal: substances which arise as end products of metabolism do not appear until some hours later. Aphids and other insects that feed on plant sap excrete a copious, clear urine as do aquatic insects, for these have no problems of water supply.

CONCLUSION

In this short review some of the ways have been mentioned in which water conservation is achieved by the terrestrial arthropods, but the subject is a very large one and many other important points could have been stressed. The vertebrates probably migrated to land by way of tropical swamps where the oxygen content of the water was low, so that the inhabitants had already become adapted to air breathing, an essential prerequisite for terrestrial life. In the case of invertebrates the course of evolution is less clear, but it is by no means impossible that the route via the littoral zone may have been the one most frequently followed.

The ancestors of the terrestrial arthropods may at first all have been restrained by behaviour mechanisms to damp environments. Later, some of their descendants, such as modern woodlice, centipedes, and millipedes, exploited still further this form of terrestrial life, while others, the insects and arachnids, acquired waterproof integuments and the other physiological and morphological mechanisms that must accompany them; but all this is pure speculation. In the absence of direct evidence comparative physiology can merely suggest the kind of course that evolution may have followed.

The general speeding up of life in a medium of abundant oxygen and negligible viscosity has resulted in increased specialization of the central nervous system, and only the size limit imposed by respiratory requirements and the necessity for growth by moulting may have prevented the arthropods from dominating the terrestrial environment to an even greater extent – possibly even to the exclusion of the vertebrates, including man.

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RESEARCH REPORT

A. W. HASLETT

NUCLEAR ENERGY

IF one concentrates on facts, the one that is most important in the peaceful development of nuclear energy is probably the British decision to proceed with a 10-year programme of nuclear power stations, based in the earlier stages on an existing design, and developing thereafter in ways not fully specified. It is a proof that, in the conditions of one country, the time for rapid application has arrived. If, on the other hand, one prefers 'ifs', then the most important is probably the extent of detailed information that may be offered by the 'have' countries to the United Nations scientific conference to be held in Geneva from 8 to 20 August; for only on rather full information can it be expected that confidence to proceed will be built. The itemized programme of the conference, which has lately been issued, provides the framework that is needed; it has been prepared on the advice of a small committee which includes representatives of Canada, the U.S.S.R., Great Britain, and the United States. It is not obvious at the time of writing whether the Geneva conference is intended to precede or to follow the attainment of agreement on the constitution of the international atomic agency which will provide the channel for assistance.

The British programme requires to be seen in the context of the American one. The circumstances are different in several ways. One is the greater need of Great Britain for an additional source of energy; another, the greater resources, including command of materials, behind the American effort; a third is the fact that, whereas electricity generation is in the hands of private enterprise in the United States, it is publicly owned in Great Britain. The apparent dependence of the

economic generation of electricity from nuclear energy on the attachment of a value to plutonium, produced as a by-product which is greater than its expected long-term value as a source of further nuclear energy. In the United States, this situation has caused the rejection of plans by industrial organizations for the combined purposes of electricity generation in competition with existing types of power stations, and of the production of plutonium for sale to the Atomic Energy Commission. The reason for this rejection has been a reluctance on the part of the Commission to enter into long-term commitments for the buying of plutonium at a price such that electricity generation from present reactors would be economic. In Great Britain, the line of thought has been that an early and expanding development of nuclear energy is a national necessity, and consequently that a relatively high value could 'reasonably' be attached to early supplies of plutonium, since its accumulation would be a prerequisite to the introduction of more efficient types of nuclear power station and to the building of 'breeder' reactors, in which fissile material would be produced than consumed. It is assumed that, while later power stations 'should show a marked improvement in efficiency ... the value of plutonium will probably fall considerably during their lifetime.' The present approach is thus justified primarily by need for energy, and secondly by confidence in the stage of rapid expansion which is expected to be reached about 1965. It is because the justification of the first stage involves the second that the need for more information arises.

Both the British and the American programmes fall into two parts - those devoted to the non-breeding and breeding types of reactors. There is a further distinction between experimental reactors, which may produce more or less of power, and reactors built in their own right for electricity generation. Only the breeding reactors are included in the 10-year British programme of stations to be built by the British Electricity Authority. On the other hand, the Atomic Energy Authority's experimental fast breeder reactor at Dounreay is provisionally expected to begin working in 1959, one year earlier than the date set

for the first production stations. There are also a number of further types of reactor, which are at the stage of 'design study' or experiment.

The United States has no equivalent of the British production programme. But, against this, there is a greater variety of power reactors built or building. Three of these are of non-breeding types. They are conventional to the extent that the maintenance of a chain reaction is dependent on the slowing up of the neutrons released in fission, and that the nuclear fuel consists of rods of uranium, slightly enriched in the fissile isotope uranium-235, and encased in a material which will meet the condition of withstanding direct contact with the fluid used to extract heat from the reactor. In each of these American designs, the fluid used is a liquid, and in two of them it is ordinary water. One, known as the 'pressurized water reactor', has been designed by the Westinghouse Electric Corporation and will be operated by the Duquesne Electric Light Company in Pittsburgh. A closed, pressurized circuit is used to extract heat from the reactor and to pass it through a heat exchanger for the boiling of water for the turbines. The other one, known as the 'experimental boiling-water reactor', is being built at the Argonne National Laboratory, and in this steam will be raised directly in the reactor. In both these designs, the water which extracts the heat will serve also as the 'moderator', by which the neutrons liberated in fission are slowed to thermal energies. The last of these three reactors will use liquid sodium to extract the heat. It is being built by North American Aviation, Inc., and will retain the traditional graphite blocks as moderator. The advantages are the greater heat conductivity possessed by a metal, and the possibility of extracting heat at a relatively high temperature without corresponding increase in working pressure. The disadvantages are that liquid sodium is an awkward material to handle, and that it absorbs more neutrons, which implies that the uranium used in the reactor must be more highly enriched in uranium-235. An economic advantage is considered to be not yet foreseeable, though it should come in time.

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development of British power reactors is based on a single, simple idea (*Science News* 32, p. 92): namely, that by fitting fins to the cartridges containing the uranium, the efficiency of heat extraction by a gas can be so far increased by breaking up the gas flow as to alter the immediate balance of advantage between gas and liquid cooling. In other words, for a given rate of heat extraction, the gas pressure necessary has been reduced significantly. The first four stations in the 10-year programme will be all of this type; two are for operation in 1960–61, and two more by 1963. Comparison of relevant figures for the first two of these stations and for the American 'pressurized water reactor' suggests that the choice has been well made. The generating capacity is expected to be about twice as great, or rather more; the cost a little more than half; and the temperature at the heat exchanger, and therefore the thermal efficiency, slightly greater (about 350° C, compared with 275° C). The comparison is unfair to the extent that the efficiency of the 'pressurized water reactor' may be improved, and that the British reactors represent a later stage in design. But in relation to these figures the statement in the white paper that this is the only type of reactor 'that is within our present technical reach in terms of design and construction in the near future' appears modestly expressed. The design of the further eight stations which are to follow has been left open. The first four may be developments of the same type. The later ones may be liquid-cooled, using either water or sodium (or an alloy of sodium and potassium); experience of the latter will have been gained in the interval from the working of the experimental Dounreay reactor.

Both for the rapid development of nuclear energy and for the efficient use of nuclear fuel there is need for the breeder type of reactor. Here there are several possibilities. The American equivalent of the Dounreay fast-neutron breeder, designed also for electricity generation, is the 'experimental breeder reactor No. 2' which is being built at the Argonne National Laboratory. The essence of this type is that a small core containing plutonium is surrounded by a blanket of either natural uranium or of uranium depleted in the fissile isotope. Plutonium is produced in the

blanket. The difficulty, when a breeder reactor is used also as a power source, is to extract heat from the small central core at the high rate necessary. For this reason, the functions of breeding and power production may have to be largely separated, although both countries are trying to combine them.

The second main possibility is the so-called homogeneous reactor, and it is not yet certain whether this will just breed or not quite. It is probably better thought of in any case rather as a non-breeder which, by economy in the use of nuclear fuel, would reduce the amount of breeding necessary for an expanding programme, than as itself contributing directly to accumulation. Its attraction lies in its simplicity on the nuclear side. In place of uranium rods, there is a circulation of nuclear fuel in solution. The reactor proper consists of a central chamber which is large enough for the usual chain reaction to proceed at the intended working temperature, and which is surrounded by a heavy-water reflector. The liquid is pumped through pipes (in which reaction ceases) to a heat exchanger, and back again to the reaction chamber. Such a system can be made self-controlling by temperature, so that the pumps can fail and be started again without difficulty arising. On the other hand, there is an obvious risk in the circulation of a highly radioactive and corrosive solution at a high temperature and pressure. The principle of this type has been established at the Oak Ridge National Laboratory, and two further reactors are planned. In one of them, the heavy-water reflector will be replaced by a blanket of thorium in which uranium-233, formed from the thorium, will accumulate. This will not necessarily amount to breeding. But breeding from thorium has been accomplished already in a fast breeder reactor. According to a recent announcement, the U.S. Atomic Energy Commission has offered to make available to private users the fissile isotopes – uranium-235, uranium-233, and plutonium – subject to accountability for the material. Presumably, a power company could, if it wished, embark on either a fast breeder or a homogeneous reactor, though this seems unlikely at the present stage.

The long-term outlook is thus fluid. Fast breeder reactors are

needed for a rapid expansion in use; but neither the fast breeder power reactor nor the homogeneous reactor are yet proved. The U.S. Atomic Energy Commission has been described by Dr. A. T. Weinberg, Research Director of the Oak Ridge laboratory, as 'hedging its bet'. There has been no hedging in Great Britain, so far as immediate decisions have been concerned. But a number of further types are being considered, as already stated, and long-term policy is no more cut-and-dried in Great Britain than in the United States.

A more specific point of progress is that the disposal - or use - of the radioactive fission products formed in any type of reactor should no longer be a major difficulty. By separating only two radio-elements from the rest the problem can be reduced to manageable proportions, while in limited quantities the two separated elements can be used. From the point of view of disposal, it has been stated that the radio-elements formed in fission can be considered as falling into three groups. At one extreme are those, the great majority, which have half-lives of less than a year. An activity with a half-life of a year would be reduced in 10 years to less than a thousandth of its original value, and in 20 years to less than a millionth. Radio-elements in this group can be looked after by storage in tanks, which can later be emptied and used again. At the other extreme, there are a few radio-elements with half-lives so long that they do not contribute importantly to total radiation. The intermediate group consists conveniently of only two elements - strontium-90, which has a half-life of about 20 years and can be used in industry for the removal of static charge from factory belts and similar non-conducting equipment on which charge can accumulate; and caesium-137, which has a half-life of 33 years and emits gamma-radiation of high enough energy (700 keV) to be useful in hospitals. A pilot plant has been built at Windscale for the separation of these two radio-elements. Chemical engineering development will be necessary for their continuous separation on a large scale. But that done, the disposal of the remaining fission products would be greatly accelerated; and, without allowance for use, the output of those separated would be only

about a ton or two a year, assuming as much electricity to be generated in nuclear power stations as is now generated from coal in Great Britain. This is the situation envisaged by 1975. Clearly a ton or two of material presents no major problem in disposal, however highly radioactive. Also the complexities of separation will be concentrated on a comparatively small number of plants; separation will not be done in nuclear power stations.

The programme of the Geneva conference makes provision for contributions on fast reactors, the chemical processing of irradiated uranium and thorium, the storage and separation of fission products, the technology of special materials needed in the more advanced types of reactors, problems of leaks, pumps, and corrosion, and the interrelation of reactor systems as a whole. It is hoped also that information will be given on certain types of reactor of which little has so far been heard. An example is the 'fluidized solid fuelled reactor'.* Also, the session on the chemical processing of irradiated fuel elements - which in plain language means the separation of plutonium - is itemized to include high-temperature processing and 'volatility processes', as well as solvent extraction. There is scope also for the discussion of safety problems in relation to the large-scale development of nuclear energy, but not those arising from military applications except so far as evidence from these is relevant. The World Health Organization has announced its intention of dealing with safety aspects, and contributions made to the conference under this heading may be referred to the latter body. It would be the appropriate body also for the sponsoring of research on long-term genetic effects, on which concern has lately been expressed.

All contributions to the conference will be sponsored by the states and specialized United Nations agencies taking part. All sessions will be open, and it is hoped that there will be at least a token admission of members of the public. No decisions will

*See the article by H. C. Pincas on p. 68 for an account of fluidized bed reactors.

be taken, or recommendations made by vote; but in the closing session there will be communications from states participating 'concerning measures for assistance to other countries in the use of atomic energy for peaceful purposes'. Considered in the normal sense as a conference, an excess both of people and contributions may be expected; and time will be given also to the uses of radio-elements in research, medicine, agriculture, and industry, which it will not be possible to discuss adequately, however relevant in a broad sense. There will be evening lectures, as at a British Association meeting, for the encouragement of wider interest. From the point of view of information, all contributions are to be circulated, and will later be published. The secretary-general is Dr Walter Whitman, Professor of Chemical Engineering at the Massachusetts Institute of Technology. The president will be Dr H. J. Bhabha (India).

A Programme of Nuclear Power (Cmd. 9389). London, H.M.S.O., February 1955.

Weinberg, A. M. 'Power Reactors'. *Scientific American*, December 1954.

'Disposing of Radioactive Waste'. *The Times*, December 31, 1954, p. 7.

RADIO-ASTRONOMY

Research by the radio section of the Cavendish Laboratory, Cambridge, has led to a possible new method for the study of the ray-like structure of the sun's corona, which by the methods of visual astronomy can be observed only during a total eclipse of the sun, and then to smaller distances from the sun's disc than the radio-astronomer can reach. The origin of the research was a suggestion by K. E. Machin and F. G. Smith in 1950 that it should be possible to observe the occultation, or masking, of a radio star when radio waves from it were suitably deviated during their passage through the corona. The first successful observations were made in 1952 on the radio star in Taurus, using the type of spaced-aerial interferometer described by W. H. Marshall in his article in *Science News* 31, and it was pointed

out at the time that some form of scattering mechanism must be involved. Further observations were made by A. Hewish in 1953, using improved equipment. The wavelengths used were 3.7 and 7.9 metres on both occasions. Variations in the apparent intensity of the radio star were found at angular distances corresponding to between 5 and 15 solar radii. The more recent and accurate measurements have made possible a theoretical discussion which suggests that the cause of scattering may lie in irregularities of electron density which 'represent an extension of structure similar to the visual coronal rays'.

A new departure of another kind in radio-astronomy has been the publication on behalf of the International Astronomical Union of a list of radio sources which are considered certain. The first such list has been issued under the name of Dr J. L. Pawsey, president of the radio-astronomy commission of the union (Commission 40), with the co-operation of workers in those centres which have chiefly contributed. It is divided into three sections. List 1 consists of 8 sources whose existence has been definitely established, for which accurate positions have been given by a number of observers, and for which optical identifications have been secured. The two most interesting are the sources in Perseus and Cygnus which have been identified with colliding galaxies. In some of these cases identification is supported by agreement between radio and optical sizes. The remaining two sections list respectively 9 sources for which the radio positions are not known so accurately but for which reasonable identifications have been suggested, and 21 sources for which there is good agreement between two or more independent radio observers. An I.A.U. number has been given to each source.

Hewish, A. *Proceedings of the Royal Society*, **A228**, 238 (1955).

Pawsey, J. L. *Astrophysical Journal*, **121**, 1 (1955).

LIVING COELACANTH

An account has been given by Professor J. Millot, Director of the Institut de Recherche Scientifique de Madagascar, of the

first observations on a living Coelacanth (*Latimeria*), the species of deep-sea fish which until 1938 was thought to have been extinct for some 60 million years. The fish was pulled in by a fisherman early at night from 255 metres depth at a distance of 1,000 metres from a jetty. A cord was passed through its mouth and gill-opening, and with difficulty the fish was 'led' to shore. Arrangements had been made in advance that a whaler should be sunk at once in shallow water to provide an emergency sea basin. The fish, which had appeared distressed on first reaching the surface, seemed to have recovered appreciably after an hour or so, and passed the rest of the night 'without any too obvious inconvenience'. After daybreak 'it became apparent that the light, and above all the sun itself, was worrying the animal very much', and several tent canvases were put over the boat. By mid-afternoon, however, the fish was dying, and it was brought into hospital for examination. It was 1.42 metres long and weighed 41 kilogrammes. 'There can be no doubt', Professor Millot states, 'that death was brought about by decompression combined with rise in temperature.' The other outstanding impression was of the extreme mobility of its pedunculate fins. Professor Millot concludes that the only hope of keeping a Coelacanth alive for a longer time with the facilities available would be to construct in advance a trellis-work cage in which the fish would normally be kept submerged at a depth of 150 to 200 metres, and would be brought to the surface only for limited times for observation.

Millot, J. (translated by D. W. Tucker). *Nature*, 175, 362 (1955).

PRODUCTION OF STABLE ISOTOPES

A reminder of the wide use made of stable isotopes in research is given by accounts published of their production at the Oak Ridge National Laboratory in the United States and at the Atomic Energy Research Establishment, Harwell, in Great Britain. In both cases the instrument of separation or enrichment is the mass spectrograph, the principle of which was described by Dr M. P. Barnett in his article in *Science News* 34;

if the incentive from research needs had been sufficient, it could have been developed as a production method independently of wartime requirements.

In the American programme, the original machines for the separation of uranium isotopes have been so far developed that two of their most important components 'now resemble only remotely those of a few years ago'; these are the ion source, in which a positive charge is imparted to the atoms or molecules to be separated, and the collectors on to which they are focused electromagnetically and separated on the basis of mass and charge. Since 1946, nearly a thousand lots of separated isotopes have been processed in the four mass spectrographs now in use, and more than 200 enriched isotopes of nearly 50 elements have been prepared. Additional elements on which work is proceeding are 5 of the rare-earth series of elements, of which suitable initial material has been lacking, and the 5 stable inert gases, where the difficulty has been in collection. Progress on both problems is reported. Rates of separation of some elements are quoted on the basis of a collector current of 100 milliamperes; but, since the actual collector current varies widely from element to element, these figures must be taken as illustrative only. They range from 26 milligrammes an hour for the two stable isotopes of lithium to 772 milligrammes an hour for four lead isotopes. The biggest group of uses has been in nuclear physics, where the ability to distinguish the reactions of one isotope from those of others is useful in many different types of experiment.

Production at Harwell, where there are two machines, both began later and is on a smaller scale. So far, 37 elements have been separated and stocks are available of most of them. A special service has been developed for the production of targets for use in nuclear bombardments. These are made by a process of secondary deposition, adapted to the element, from already separated stocks. In this way targets are produced within a few days to meet the varied requirements of users. There is a further advantage that the final deposition starts from a definite amount of material of known purity. Both at Oak Ridge and at Harwell some of the separated stable isotopes are used also in the produc-

tion of radioactive isotopes from them by irradiation in a nuclear reactor.

Keim, C. P. *Nature*, **175**, 98 (1955).

Allen, W. D., Dawton, R. H., Smith, M. L., and Thonemann, P. C., *Nature*, **175**, 98 (1955).

TRACKING THAMES MUD

The application of the radioactive tracer method to follow the later history of Thames mud dredged from the Port of London area, and mostly dumped at sea beyond the estuary, has been reported by the Hydraulics Research Station, Wallingford, of the Department of Scientific and Industrial Research. The problem was to provide a tracer material which would reproduce as closely as possible the movement of natural mud particles. Glass, ground to a similar average size, was found to be suitable, and a glass incorporating scandium oxide was prepared and was irradiated in a nuclear reactor at the Atomic Energy Research Establishment, Harwell. Radioactive scandium-46 was thus formed. The treated and ground glass was then mixed with river mud, and deposited in mid-channel at Gravesend at river-bed level, using a drum which could be opened by pulling a release wire. This was done at slack water. A search for radioactivity on the river bed was carried out during the next several days, using Geiger counters in a waterproof container. The quantity of material released was insufficient to enable its movements to be followed for more than a few tidal cycles. During this period, some of it was carried at least two or three miles upstream. It was found also that the tracer material became so widely dispersed and diluted with normal Thames mud that there could be no danger even if dredging operations were carried out later in the area. The method is thus thought to be practical and safe for the following of river-bed material, if used in a larger experiment.

ARTIFICIAL DIAMONDS

Small artificial diamonds of industrial quality have been made by F. P. Bundy and others in the General Electric Research

Laboratories, Schenectady, New York, by a process in which unstated compounds of carbon have been subjected for varying periods to different combinations of high temperature and pressure. The greatest pressure used (100,000 atmospheres) was the same as that used by Professor P. W. Bridgman in his measurements at Harvard University on the physical properties of compressed solids. This was his limit for useful experiments. The highest pressure which he attained – without ability to make measurements – was about 420,000 atmospheres. Under this pressure, but without special heating, he calculated that graphite should be transformed into diamond, but reported that this did not happen. In one combination of conditions used in the General Electric laboratories, heating was continued for 16 hours at a temperature above 2,800° C and under a pressure of slightly over 50,000 atmospheres. A crystal nearly $\frac{1}{16}$ inch long was obtained. Under other conditions, diamonds have been produced in a matter of minutes. They have been identified by X-ray and chemical examination, and have passed hardness tests. It is stated that no thought has been given to mass production, and that no attempt has been made to produce diamonds of gem quality.

Science, **121**, 228 (1955).

CORRESPONDENCE

FINDING OUT ABOUT FUNDAMENTAL PARTICLES

Before the publication of *Science News* 35, but too late for correction, I found a mistake in my article. The sentence beginning two lines from the bottom on p. 29 should have read: 'The spin of the π -meson is zero, and so the spin of the τ -meson is zero also.' Correspondingly, in Table I, on p. 26, the spins of these particles should have been given as zero. These corrections are important. There are also two errors of typography: in the same table, the symbol Ω^+ should have been used (under 'Decay scheme') for the omega particle, and Λ^0 (under 'Symbol') for the neutral lambda particle.

R. M. TENNENT

Department of Physics,
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THE RIVERS WEAVER AND DANE

In *Science News* 33, Inset 22, you show an aerial photograph of the Dane behind Shurlach. This you have titled, incorrectly, 'The River Weaver at Northwich'. I am alarmed that a mistaken impression of the Weaver may thus be perpetuated. We have been slowly but steadily building our Whit-Monday Regatta up till it has now somewhat more than a purely local reputation. One of the facilities on which we rely to attract crews from far afield is an extremely fine mile-long stretch of water, which, although not entirely straight, has two equal and opposite bends of a mere 9° either way. You can understand, therefore, that a would-be competitor seeing the photograph authoritatively entitled 'The River Weaver at Northwich' would be hard put to it to reconcile the convoluted ditch illustrated with the finest course in the North of England to which we were hoping to attract his crew.

A. S. IRVINE, Chairman,
Northwich Regatta.

Dr Dury writes:

The caption should have read: 'The Weaver valley at Northwich. One valley bend appears in the foreground, with some ten river meanders, mostly in the tributary Dane, and cut-offs in the present cycle'. The main river, at the top of the photograph, is unfortunately much less apparent in the reproduction than in the original.

SOME BOOKS RECEIVED

CULTURE AND HUMAN FERTILITY by Frank Lorimer and others (Paris, Unesco; U.K. agents, H.M.S.O., 1954), pp. 510, 25s.

Professor Lorimer's approach to human populations has been from the side of sociology, and the present study of 'the relation of cultural conditions to fertility in non-industrial and transitional societies' bears the stamp of this interest. A general review and theoretical discussion by Professor Lorimer is followed by accounts of investigations in the Gold Coast by Professor Meyer Fortes and Professor K. A. Busia; in Buganda and Buhaya (populations of relatively low birth-rate) in East Africa, by Dr Audrey I. Richards and Mrs Priscilla Reining; in Brazil, where 'the predominant influence is the biological factor', by Professor G. Mortara, who summarizes a series of studies carried out by the Instituto Brasileiro de Geografia e Estatística. Although every contribution is individually of interest, the investigations described vary in scope, emphasis, and degree of intensiveness, and not even the replacement of a would-be directing committee by Professor Lorimer has given more than relative unity to the whole.

FACT, FICTION, AND FORECAST by Nelson Goodman (University of London, The Athlone Press, 1954), pp. 126, 15s.

Compared with Professor Braithwaite's *Scientific Explanation*, Professor Nelson Goodman is concerned rather with induction as a process than with deductive systems, with the philosophy-as-such which underlies science than with explaining to scientists what they do. Beginning with 'counterfactual conditionals' (in which the antecedent is false), he develops the suggestion that 'the line between valid and invalid predictions . . . is drawn upon the basis of how the world is and has been described and anticipated in words'. Though the working scientist may see little help in this hoped-for solution from use of language, he might learn from the author that the word 'essentially' followed by 'the same' or 'similar' was usually a fair indication of a question begged.

Some Books Received

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THE HUMAN ANIMAL by Weston La Barre (University of Chicago Press; U.K. agents, Cambridge University Press, 1955), pp. 372, 45s.

Probably the most useful comment on Professor La Barre's book is that those nurtured in the orthodox susceptibilities of biologists should not be put off by his early definitions of living organisms and evolution. As an anthropologist who draws on psychology and sociology, but little on physiology, he gives a picture of enough of man to be interesting and includes enough ideas of his own to make others think.

THE PILTDOWN FORGERY by J. S. Weiner (Oxford University Press, Geoffrey Cumberlege, 1955), pp. 214, 12s 6d.

In contrast with other accounts, Dr Weiner (with more late corrections than he can have wished) here concentrates on the human problem – one of the most intriguing raised by science – but concludes that the weight of circumstantial evidence falls short of proof.

VARIABLE STARS AND GALACTIC STRUCTURE by Cecilia Payne-Gaposchkin (University of London, The Athlone Press, 1954), pp. 116, 18s.

For anyone who has become interested in the interrelated problems of variable stars, obscuring material, star populations, and the geography and evolution of our own Galaxy and of the extragalactic nebulae, this book collects much material that belies the precision with which arguments based on it have often been presented; concentration is required, but reading is not seriously difficult. In a final survey the information given is described as for the most part vague in detail and provisional, although on main outlines certainty is growing.

EXPLAINING THE ATOM by Selig Hecht revised with four additional chapters by E. Rabinowitch (London: Victor Gollancz Ltd, 1955), pp. 224, 12s 6d.

Dr Rabinowitch has done a good job in bringing up to date the quiet and simple account written by Professor Hecht in 1946, though with not quite the distinction of the original. Professor Hecht's concluding paragraphs, which remain valid and moving, are retained.

The book has not before been published in a British edition, though it deserved to be.

X-RAY DIFFRACTION BY POLYCRYSTALLINE MATERIALS edited by H. S. Peiser, H. P. Rooksby, and A. J. C. Wilson (London: the Institute of Physics, 1955), pp. 725, 63s.

A magnificently produced and useful book which many students will wish that their libraries should acquire for them. Its division into three parts – experimental methods, interpretation, and essays on results obtained in different fields of use – makes it possible to gain a quicker and more realistic view of one of the two main divisions of X-ray analysis than might be expected from the size of the book and the necessity for some thirty contributors.

APPLIED MASS SPECTROMETRY: REPORT OF A CONFERENCE, 1953 (London: The Institute of Petroleum, 1954), pp. 333, 50s.

This report of a conference held in London in October 1953 is listed in completion of the further reading suggested by M. P. Barnett in his article in *Science News* 34; conversely, Barnett's article could be used now as an introduction to the report.

THE HYDROGEN BOMB by James Shepley and Clay Blair Jr. (London: Jarrolds, 1955), pp. 216, 12s 6d.

A book which leaves a taste in the mouth. See p. 95 for some statements relevant to the setting off of thermonuclear reactions; *passim* for anti-Oppenheimer prejudice.

ABOUT OUR CONTRIBUTORS

J. L. CLOUDSLEY-THOMPSON is a lecturer in zoology at King's College London. Born in Kashmir in 1921, he was educated at Marlborough and Pembroke College, Cambridge. He is honorary secretary to the Biological Council, and has written previously in *Science News* 23 and 28.

L. R. B. ELTON studied mathematics at Christ's College, Cambridge, during the war, and took his Ph.D. in quantum mechanics at University College London in 1950. He is now a lecturer in theoretical physics at King's College London.

F. A. FOX graduated in metallurgy at Birmingham and has done research in various fields of physical metallurgy (D.Sc., 1947). He is now working in Salisbury, South Australia, directing research in physics and chemistry in the Australian Defence Scientific Service.

J. O. LAWS was at Cambridge University and Guy's Hospital, graduating in 1943. From 1946 to 1951 he was Gull Student in the Pathology Department at Guy's Hospital Medical School, working on problems of immunology and cancer research. After obtaining his Ph.D., he spent a year at the Institut du Radium in Paris before taking up his present position as Research Fellow in the Department of Experimental Pathology and Cancer Research at Leeds University.

A. M. MAYER, aged 28, is a graduate of University College London, who turned from chemistry (B.Sc.) to plant physiology (Ph.D.). Since 1950 he has worked in the Botany Department, Hebrew University.

H. C. PINCAS was born in Russia, educated in Germany, graduating in the University of Berlin (Ph.D.), and worked in various industries and countries in Europe and England before settling, in 1938, in Australia, where he is doing research work on coal in the New South Wales University of Technology.

I. S. WILLIAMS read physics at King's College London and, after graduation in 1950, obtained a Medical Research Council post-

