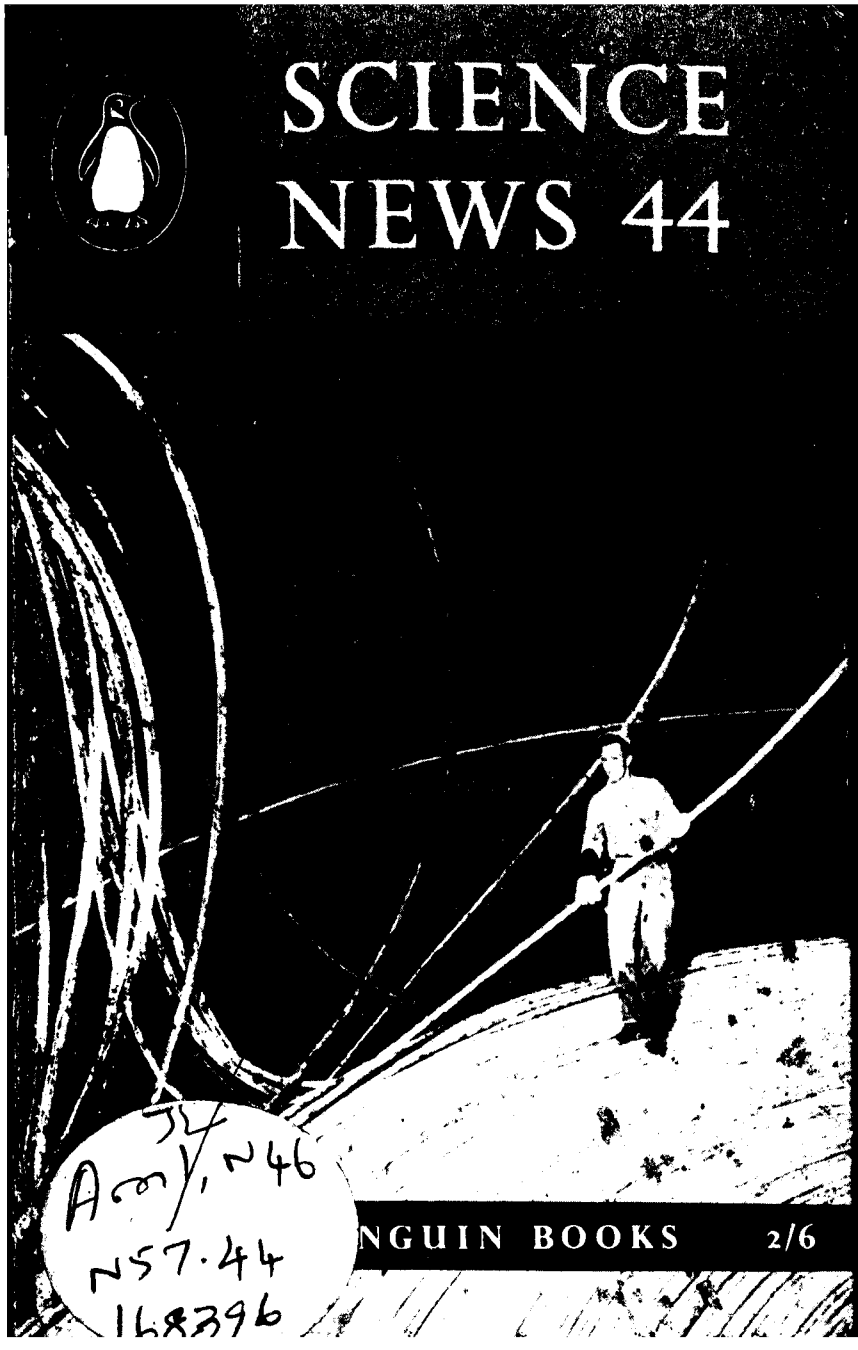
SCIENCE NEWS 44



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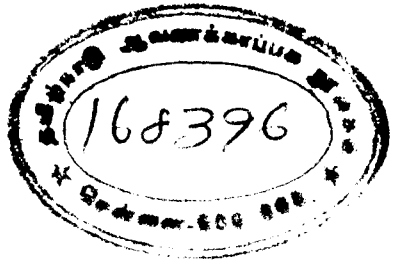

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CONTENTS

A NEW OUTLOOK IN INORGANIC CHEMISTRY

R. S. NYHOLM

THE outstanding developments during the past decade in the production of atomic energy by the use of nuclear reactions have focused attention upon inorganic chemistry, a subject that had for many years previously been seriously neglected. Among these chemical developments one thinks at once of the production of the hitherto unknown *trans-uranic* elements (actinides), of which nine are now known, one of the most spectacular achievements ever recorded in preparative inorganic chemistry. Similarly, nuclear fission has led to the filling in of the gaps* in the Periodic Table of 1939, i.e., elements 43 (Technetium), 61 (Prometheum), 85 (Astatine), and 87 (Francium). The separation and identification of these elements, however, has involved a number of new experimental techniques, such as the use of ion-exchange resins, as well as the application of the whole armoury of physical chemistry. This work has served to draw attention to the remarkable change that has been taking place in the approach to inorganic chemistry during the past two decades. This new approach has transformed the 'old inorganic chemistry' of twenty-five years ago into a new and exciting subject with a real sense of purpose; one that possesses hitherto unrecognized unity within itself and with chemistry as a whole. Furthermore, as will be shown, it provides great opportunities for those interested in chemical research in its widest sense as embracing all techniques and branches of chemistry. In this article the reasons for this new outlook and the events that have led up to it will be

*The author cannot resist uttering a plea at this point for the long overdue filling in of these gaps on many of the old Periodic Tables that hang on the walls of schools, colleges, and universities.

discussed. Then a picture of the subject at the present time will be given, and finally an attempt will be made to indicate the work that is going on at some of the rapidly expanding frontiers.

THE DEVELOPMENT OF MODERN INORGANIC CHEMISTRY

Those of us who were familiar with the state of inorganic chemistry in schools and universities twenty to thirty years ago will recall that at that time it was widely regarded as the least interesting part of the chemistry course. Usually, it was taught almost entirely at school and in the early years of the undergraduate course and then chiefly as a collection of largely unconnected facts. On the whole one tended to conclude that, apart from some relationships dependent upon the Periodic Table, there was no system in inorganic chemistry comparable with that to be found in organic chemistry, and none of the rigour and logic that characterized physical chemistry. It was widely believed that the opportunities for research in inorganic chemistry were few, and that in any case the problems were dull and uninspiring; as a result, relatively few people specialized in this subject. Unhappily the effect of this neglect became apparent during the 1939 war and in the post-war years, when chemists with a sound knowledge of inorganic chemistry were required for the development of atomic energy projects. So long as inorganic chemistry is regarded, as in years gone by, as consisting simply of the preparation and analysis of elements and compounds, its lack of appeal is only to be expected. But this stage is now past and in keeping with the new broad outlook in the subject we can define inorganic chemistry today as the integrated study of the formation, composition, structure, and reactions of the chemical elements and their compounds excepting most of those of carbon. This is certainly an all-embracing definition, a point that must be made at the outset because modern inorganic chemistry overlaps in all directions the other two traditional branches of the subject – organic and physical chemistry – not forgetting biological chemistry, the study of the chemistry of compounds of importance in biology. Physical and inorganic

chemistry are, of course, closely related, but if one seeks differences between them, the former is concerned more with the development of theories and physical methods, the latter more with their application to the study of the structure and properties of chemical compounds.

The factors primarily responsible for the modern forward-looking spirit in inorganic chemistry are two developments in other fields, one of which gives it a new sense of purpose, and the other, new tools with which to achieve this purpose. The first of these developments is the growth of the theoretical techniques of quantum mechanics to an extent permitting widespread chemical application. This has already proceeded far enough to show the unity of inorganic chemistry even though quantitative experimental investigations are required to reveal its full significance. The second external development consists of those new optical, electrical, and magnetic techniques of physical measurement, by which structure can be investigated in the physical terms demanded by the now fully accepted electrical nature of matter. For a full appreciation of the way in which these advances have affected the development of inorganic chemistry a brief survey of the history of the subject over the past century is helpful.

During the 19th century inorganic chemistry, initially the chemistry of metals, minerals, and rocks, developed side by side with organic chemistry, the chemistry of compounds initially believed to be formed only in living organisms but later defined as the chemistry of carbon compounds. The main work carried out in inorganic chemistry was concerned with the preparation of new compounds and the development of methods of analysis. Vast numbers of new compounds were prepared and analysed. In addition, important work was carried out on the determination of atomic weights. The year 1887 may be accepted as the date of emergence of physical chemistry as yet another branch of the subject, and many research workers were attracted to physical chemistry because it offered an exactness that was lacking in inorganic chemistry. In the meantime, stimulated by the postulate of the four-covalent tetrahedral carbon atom

developed simultaneously by van't Hoff and Le Bel in 1874, organic chemistry developed into a system in which it was possible by classical methods to determine structure. Denied the techniques needed for the investigating of the shape of molecules, inorganic chemistry lagged behind. Thus we find that by this time organic chemistry, because of its system, and physical chemistry, because of its exactness, were steadily attracting workers from the more empirical and much less integrated fields of inorganic chemistry. It has been said that facts give a science subject its substance, but it is the integrating theory that provides its strength. There was no lack of substance in inorganic chemistry in the late 19th century, but it was sadly lacking in cohesion. It is just this development of integrating theory that has been mainly responsible for the revival of inorganic chemistry.

At the turn of the century Alfred Werner did much to put some order into inorganic chemistry. Werner is specially remembered for his studies on the apparently anomalous inorganic addition compounds, many of which had been described during the preceding half-century. These addition compounds, which have fixed and definite composition, can be formed by the union of two or more previously stable and saturated molecules. Thus, white anhydrous copper sulphate unites with four molecules of ammonia gas to yield the intensely blue tetrammine cupric sulphate, i.e.,

$$CuSO_4 + 4NH_3 \longrightarrow CuSO_4.4NH_3$$

In this substance the four ammonia groups are arranged about the copper atom in the form of a square as shown in Figure 1. These addition compounds are now known as complex or co-ordination compounds and the groups that attach themselves are called 'ligands'. Werner put forward the first satisfying theory for the structure of these compounds. His theory involved the idea of a definite *co-ordination* number of a metal with the ligands arranged about it according to definite geometrical patterns of which the octahedral, the tetrahedral, and the square planar arrangements (see Figures 2, 3, 4) are most important.

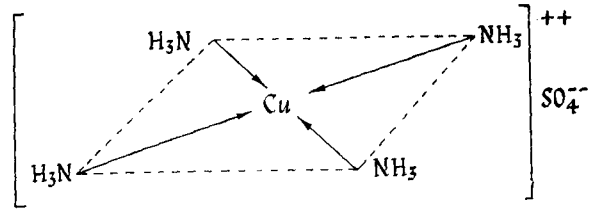


Fig. 1. Structure of tetrammine cupric sulphate.

For the first time this theory enabled many of the apparently conflicting data to be tied together. Paradoxically, however, Werner answered so many questions that he left many people with the impression that in inorganic chemistry there were but few advances to be made. However it is a sobering thought to recall that no satisfactory explanation of the nature of chemical combination, or of valency, had then been advanced and our knowledge of the shape of molecules was still limited to two or three common shapes.

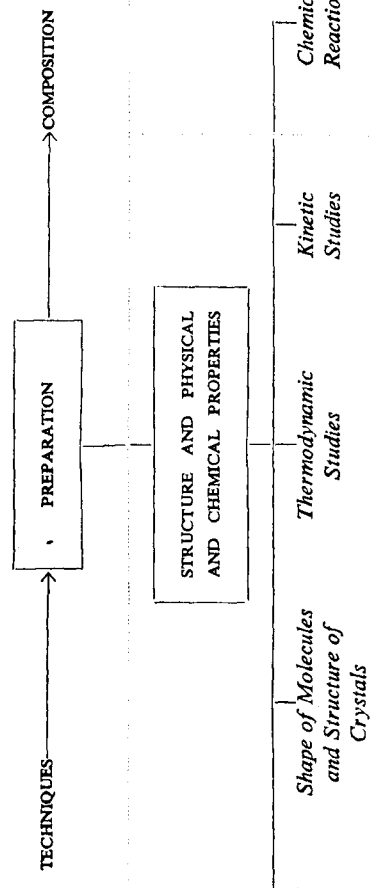
From the first world war onwards inorganic chemistry tended to be neglected. There were a few laboratories where the study of complex compounds continued, and some important work was done, but as the workers were limited in the number of techniques that were of value to them, their conclusions as to structure were, in most cases, only tentative. Excellent preparative work was carried out on various subjects, such as the boron hydrides, but in comparison with organic and physical chemistry the amount of such work was small. The development of X-ray crystallography was a great stimulus, but unfortunately the technique had to be restricted to simple compounds and progress was relatively slow. After the first world war, however, ideas on valency gradually began to take firm shape. It is important to stress the need for developments in valency theory and stereochemistry, because without an advance in these, any hope for an understanding of much of the factual matter of inorganic chemistry itself could not be expected. A great step forward occurred in 1927 with the publication of N. V. Sidgwick's

outstanding monograph, *The Electronic Theory of Valency*. Shortly after this quantum mechanics was applied to problems of chemical combination. Of special importance to inorganic chemistry was the work of Linus Pauling. By the time we reach the middle thirties the new valency theory was being applied to a wide variety of compounds. Quantum mechanics not only provided an explanation of how atoms combined, it also led on to some understanding of the strength of bonds and offered an explanation of the orientation of the bonds in space. Equally important was the fact that quantum mechanics gave a sense of purpose to many physical measurements which previously were of much more limited value in deciding the structure of chemical compounds.

It is my belief that the impact of quantum mechanics and of modern physical methods of attack are the main reasons for the new outlook of inorganic chemistry which has led to the present period of rapid growth. Of course it has often been pointed out that many of the early hopes of quantum mechanics have not been realized. Sir Harold Hartley quotes Lord Rutherford as having said in 1919: 'What's the use of going back to chemistry when Bohr will soon be able to calculate anything you can find out?' Similar optimism prevailed in 1929 when the new quantum mechanics was applied to valency, but it soon became apparent that in real molecules great mathematical difficulties are involved in obtaining a rigorous solution. As a result, various methods of approximation have to be adopted. The most fruitful results have accrued from quantum mechanics when the theoretical workers are allied with the more experimental investigators and inorganic chemistry has benefited most effectively from this kind of co-operation.

The Nature of Modern Inorganic Chemistry

The difference between the old and the new approaches to inorganic chemistry can be shown by a table which gives a simplified genealogy of an inorganic chemical investigation. The dotted line indicates roughly the borderline between the old and the new outlook; to make a new compound, to establish its



Nature of the Bond	*Shape of Molecules and Structure of Crystals*	*Thermodynamic Studies*	*Kinetic Studies*	*Chemical Reactions*
Valency of the atom; whether ionic or covalent; bond order if covalent; bond strength.	Arrangement in space of covalent bonds around central atom, and packing of ions or molecules to form crystals.	Heat of formation and dissociation; equilibrium investigations; feasibility of reactions at various temperatures.	Rates of formation and decompositions.	Reaction products and steric course of reaction.

TABLE 1: *Inorganic Chemical Investigation*

composition by analysis, and to study its behaviour were once thought to be enough. To go on from this point to determine the structure and other physical and chemical properties is the feature that distinguishes the modern inorganic chemist. It must be emphasized however that, in the past, it was often not any lack of curiosity, enthusiasm, or indeed ability, that prevented inorganic workers from going on – it was simply that the necessary valency theory was undeveloped, and the physical methods of investigation were either not available or the results from them could not be interpreted owing to the lack of a suitable theory. However, the pattern of the older kind of inorganic chemistry as a descriptive subject concerned mainly with analysis and reactions which was forced upon chemists fifty years ago is still too evident in many text-books and in much of our present-day teaching long after the necessity for such a limited approach to the subject has disappeared. The other feature emphasized by this table is the way in which the frontiers between the various branches of chemistry are being broken down. Organic groups are often attached to metal atoms today if the ensuing properties make the compound more simple to investigate and, of course, the whole of physical chemistry is interwoven into the fabric of inorganic chemistry.

It has become obvious that the central theme of inorganic chemistry, indeed the whole of modern chemistry, is a valency theory, the theory of the binding between metal atoms. Of course valency theory depends, in turn, upon the electrical nature of the atom. The basic idea of Bohr that an atom has a positively charged nucleus and is surrounded by negatively charged electrons revolving in circles and ellipses has been replaced by electrons in *orbitals*. These are conveniently represented by regions of space in which there is the maximum probability of finding electrons. According as an electron has a subsidiary quantum number of 0, 1, 2, 3, etc., it is said to occupy an *s*, a *p*, a *d*, or an *f* orbital. The *s* orbital can be likened to the region of space occupied by the skin of a very thick-skinned orange – it is non-directional so far as an approaching atom is concerned. The three *p* orbitals are shaped like dumb-bells mutually at right

angles. It is the directional properties of these that explains the *V* shape of the water molecule, and the pyramidal shape of ammonia, to give but two examples in aspects of valency theory. These have been discussed in *Science News 20* and space precludes other than a brief summary here.

When two atoms *A* and *B* (e.g., atoms of sodium (Na) and chlorine (Cl) or an atom of hydrogen (H) and an atom of iodine (I)) combine to form a compound *AB* we recognize two extreme kinds of bonds, *electrostatic* and *covalent*, with gradations between them. If the compound exists in the form of cations (positive ions) and anions (negative ions), e.g., Na^+Cl^-, the binding is electrostatic, but when there is a sharing of electron pairs between the two combined atoms covalent binding occurs. In the first example the three-dimensional arrangement of the ions about one another is decided by the relative numbers of each, the relative sizes of each, and the ease of deformation of each by electrostatic forces. The covalent bond results from the overlap of atomic orbitals on *A* and *B*. These orbitals each containing an unpaired electron have various shapes and this of course immediately leads us to the idea of *directed valency*.

In Figure 2 the various kinds of valency forces are summarized.

At this point it is valuable to say a little about the shape of molecules, since here is a good example of the way in which facts of inorganic chemistry may be tied together. In covalent molecules the shape depends upon the way in which both the bonding and non-bonding pairs of electrons in the valency shell of each atom are arranged in space. There are several 'models' used by quantum mechanists to enable this disposition in space to be predicted. The most frequently used method is the *valence bond method* which assumes that in a molecule the bonding electron pairs will occupy *orbitals* that are, to a first approximation, directed in space in the same way as in a free atom. In practice there are deviations from the ideal shape which can be understood by assuming that other possible configurations contribute to the final shape (which is then a *hybrid*). Of the other methods the simplest is the so-called *electron pair repulsion model*. To

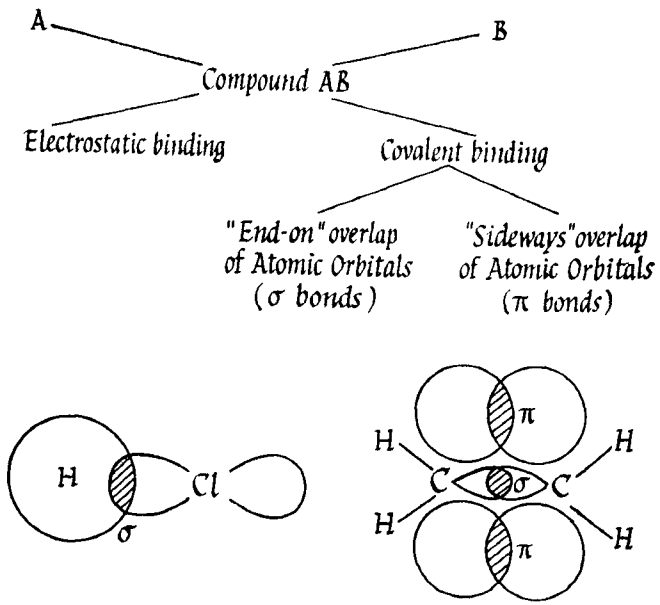


Fig. 2. Bonding due to overlap of atomic orbitals – end on to give σ bonds and sideways for π bonds.

obtain the shape of a molecule (other than those of the transition metals) from this theory one adopts the following rules:
(*a*) The shape of a molecule is decided by both the non-bonding (lone) pairs of electrons and the internuclear bonding pairs of electrons in the valency shell of an atom.
(*b*) These electron pairs dispose themselves in space in such a way as to ensure that the repulsion forces between them are a minimum.

Table II shows how well this accounts for the shape of most simple molecules. For this simple picture we have ignored any double or triple bonding pairs of electrons, but this apparent omission does not seriously affect the arguments, because, to a first approximation, the pairs of electrons between the nuclei

Total Number of Electron Pairs in Valency Shell	*Expected Arrangement of Electron Pairs in Space*	*Number of Inter-nuclear (σ) Bonding Pairs*	*Number of Lone Pairs (Non-Bonding)*	*Example*	*Shape*
2	Linear	2	0	$HgCl_2$	Linear
3	Triangular Plane	3	0	BCl_3	Triangular Plane
		2	1	$SnCl_2$(gas)	V-shaped
4	Regular Tetrahedron	4	0	CH_4	Regular Tetrahedron
		3	1	NH_3	Pyramidal
		2	2	H_2O	V-shaped
5	Trigonal Bipyramid	5	0	PCl_5	Trigonal Bipyramid
		4	1	$TeCl_4$*	Irregular Tetrahedron
		3	2	ClF_3*	T-shaped
		2	3	$[ICl_2]^-$*	Linear
6	Regular Octahedron	6	0	SF_6	Regular Octahedron
		5	1	IF_5	Square Pyramid
		4	2	$[ICl_4]^-$*	Square Planar
7	Pentagonal Bipyramid	7	0	IF_7	Pentagonal Bipyramid

* Ambiguity exists as to the location of the lone pair(s) in these molecules.

TABLE II: *Shapes of Covalent Molecules from Repulsion of Electron Pairs*

decide the shape. It must be stressed also that some deviations from the ideal shapes shown commonly occur as a consequence of multiple bonding and the attachment of different groups amongst other reasons. Furthermore, ambiguity exists in the example shown owing to there being two or more different ways in which lone pairs can be put into the arrangement. In spite of this the table does show that simple valency theory can lead to a great simplifying of structural inorganic chemistry.

Modern valency theory also enables us to understand why some bonds are weak and others strong, why some compounds are reactive and others are not, and above all that it is complementary to physical methods of investigation. In Table IV we give a list of some of these physical methods, most of which are to be found in use in a large modern school of chemistry. Is it any wonder that a modern chemist must receive in his course a sound training in physics and mathematics?

Research in Inorganic Chemistry Today

In the foregoing sections the way in which a new outlook in inorganic chemistry has developed has been traced, and it has been suggested that a more purposeful approach to research has resulted from the application of quantum mechanics and physical methods of attack. This enables us to indicate the main frontiers of advance in inorganic chemistry today and to look at some of the work that is being carried out. The following summary illustrates these major fields of development.

(*a*) Study of complex (co-ordination) compounds: their characterization, stability, stereo-chemistry, and reactions.

(*b*) Stabilization of new or previously rare valency states of metals.

(c) The study of chemistry of newly discovered elements, e.g., Pm, Tc, At, Fr, and Actinides.

(*d*) Organo-metal chemistry, e.g., cyclopentadienyl compounds of transition metals and the chemistry of metal carbonyls.

(*e*) Chemistry of lighter elements (first two rows of Periodic Table), e.g., boron hydrides, sulphur-nitrogen compounds.

(*f*) Study of the stereo-chemistry of inorganic compounds including the preparation of compounds of less common co-ordination numbers (e.g., 3, 5, 7) and the study of their structure.

(*g*) Reaction mechanisms in inorganic chemistry.

(*h*) Equilibrium studies on inorganic compounds. Thermochemical measurements generally.

(*i*) Biological chemistry of metal ions.

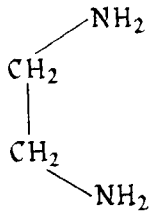
(*j*) The chemistry of the solid state.

Complex or Co-ordinated Compounds

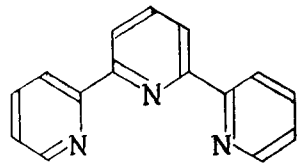
The nature of complex compounds was referred to above, the essential idea being that bound to a metal ion is a definite number of groups (now called 'ligands', i.e., something that is attached) which are oriented in space in a definite way. For a long time these co-ordination compounds were regarded as a special class of inorganic substances, but during the past few years it has been gradually realized that, especially in aqueous solution, they form the central theme of inorganic chemistry.

Research into the nature of complex compounds of metals is extremely important for developments in other branches of chemistry, in industry, and in medicine. Of special interest are the so-called 'chelate' groups which are able to attach themselves at more than one position around the metal atom. The following are examples of chelate groups which can occupy 2, 3, 4, and 6 positions around a metal atom. The term 'chelate group' is derived from the Greek 'chelos', a crab's claw, emphasizing that these compounds attach themselves by more than one 'claw'. The term 'poly-dentate group' is also used to suggest many 'teeth' gripping the metal atom.

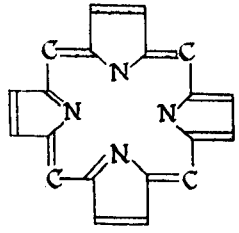
The hexadentate group, mentioned above, is of special importance industrially because it can bind metal ions very strongly and prevent them from being precipitated at inconvenient stages of a process. The quadridentate group is of importance in biology because, with an iron atom attached to the four nitrogen atoms, it becomes the nucleus of haemoglobin – the catalyst that enables us to use oxygen at room temperature for combustion and the production of energy in our bodies. Finally, the study of complex compounds is important in connexion with



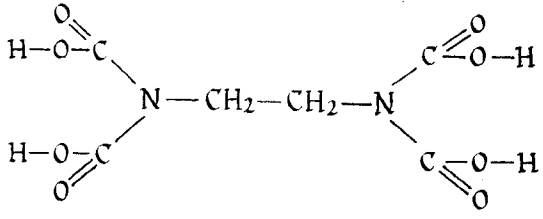
(a) Bidentate chelate group (Ethylene-diamine)



(b) Tridentate chelate group (Tripyridyl)



(c) Quadridentate chelate group



(d) Hexadentate chelate group Ethylenediamine tetra-acetic acid

Fig. 3. Examples of various poly-dentate groups.

the problems encountered in the separation of metal ions from aqueous solutions, for these usually resolve into a study of the nature of the metal complexes present.

New Valency States

Compared with thirty years ago, the number of established valency states of the transition metals is alarming – or exciting – according as one has to learn them for an examination or not. For example, thirty years ago valencies for iron of other than two (e.g., $FeCl_2$) or three (e.g., $FeCl_3$) were considered doubtful or non-existent. In the carbonyl $Fe(CO)_5$, the obvious valency state of zero for iron was regarded as 'peculiar'. Today iron is freely accepted as having valencies of O, I, II, III, IV, V, and VI. The reason for the recent isolation of so many new valency states is primarily that we understand a good deal of how the stabilization of (so-called) unusual valency states occurs. Broadly speaking, metal atoms in compounds do not like to acquire a large positive charge (i.e., greater than $+1e$ units) and are even less likely to become negatively charged (i.e., greater than $-1e$ units). Now if the new compound is an electrostatic assembly of positive and negative ions (e.g., the compound Al^+F^-) its stability will depend upon whether or not the energy evolved when the cations and anions come together to form the crystalline lattice is large enough to provide the work needed to convert the solid metal to gaseous atoms and then to remove the necessary electrons to yield a positive ion. Knowing the sizes of ions, the ionization potentials of metals, the electron affinities of anions, the heats of vaporization of metals, and heats of dissociation of halogen molecules into atoms, it is possible to predict whether or not various metal halides will be stable at room temperature. Data concerning electrostatic compounds of most of the elements in the Periodic Table have been studied in this way, because of their relevance to atomic energy projects, and the conditions for the existence of various compounds were indicated. This work shows, for example, that whereas aluminium monofluoride (AlF) is a stable species in the gas phase at 1,200° C, it is unstable at room temperature and decomposes

according to the equation $3AlF \longrightarrow 2Al + AlF_3$. This offers, incidentally, a possible method for making pure aluminium. One might heat AlF_3 with excess aluminium and rapidly cool the gaseous AlF formed yielding pure aluminium and the AlF_3 again. Similarly it can be shown that the simple fluorides of tervalent nickel (NiF_3) and copper CuF_3) are thermodynamically unstable at room temperatures. For obtaining covalent compounds of metals in *low* valency states, e.g., zero valency, the main problem appears to be one of avoiding a high negative charge on the metal atom. Thus in the following structure for nickel tetracarbonyl the apparent change on the nickel atom,

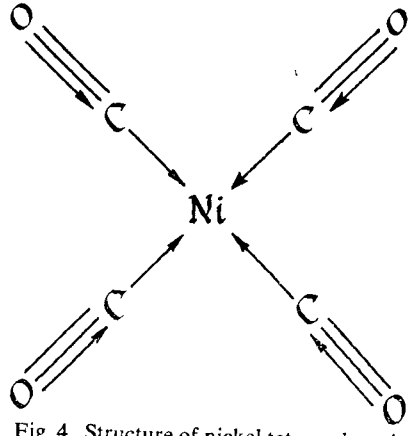


Fig. 4. Structure of nickel tetracarbonyl.

if equal sharing of the four pairs of electrons between the carbon atoms and the nickle atom is assumed, is -4 units. In fact, we believe that some double bonding occurs whereby the nickel atom feeds back other *d* electrons to each carbon atom, thereby decreasing the charge upon the metal atom and enhancing the stability of the compound. The application of these ideas has led to the use of other ligands, e.g., the isocyanide RNC yielding in turn a wealth of new and interesting compounds in which other metal atoms are zero-valent.

Thermodynamics	*Electrical*	*Magnetic*	*Spectroscopic*	*Diffraction of Rays and Particles*	*Radioactive and Heavy Isotopes*
Molecular weight measurements using (*a*) change in vapour pressure; (*b*) boiling point; or (*c*) freezing point, of a solvent on adding a solute. For large molecular weights osmotic pressure studies are useful.	(*a*) *Conductivity.* For determining if a substance is an electrolyte and the number and nature of the ions, (*b*) *E.M.F.* For determining presence of various ions and their concentration, e.g. silver ions using silver electrode. (*c*) *Capacity.* For determining presence of electric dipoles.	*Susceptibility Measurements* give number of unpaired electrons in atoms or molecules leading to knowledge about valency bond type and shape of molecule.	Methods which involve *emission* or *absorption* of *electro-magnetic radiation* of various wavelengths.	The study of the *diffraction* of (*a*) X-rays (*b*) electrons and neutrons by solids, liquids, and gases gives data concerning the position of the atoms in the compound.	Using *electronic* counters and *mass spectrographs* respectively these isotopes enable one to study mechanism of chemical reactions.

TABLE III: *Physical Methods for Studying Structure*

Rates of Inorganic Reactions of Equilibria

Finally, I want to say something about two important fields of research with special reference to complex compounds. When a chemical reaction occurs one wants to know, amongst other things, the rate at which the starting substances react and the extent to which the reaction proceeds. These two are not necessarily related even though they are often confused. Rate studies fall under the general heading of kinetics, but studies of the extent to which a reaction proceeds deal with equilibria, the thermodynamics of the reaction. Some examples are helpful. Both tin tetrachloride ($SnCl_4$) and carbon tetrachloride (CCl_4) will, given time enough, react with water at room temperatures as shown by the following chemical equations

(1) $SnCl_4 + 3H_2O \longrightarrow H_2SiO_3 + 4HCl$

(2) $CCl_4 + 3H_2O \longrightarrow H_2CO_3 + 4HCl$

Reaction (1) takes seconds to reach equilibrium whereas reaction (2) takes many years! As an example of equilibria we might compare the mixing together of hydrogen (H_2) and chlorine (Cl_2), and hydrogen and iodine (I_2) at 500° C. The former react to yield hydrogen chloride (HCl) in almost quantitative yield, whereas H_2 and I_2 combine only to the extent of about 75 per cent.

The study of the equilibria of complex compounds in aqueous solution is now a very rapidly expanding field of research. The measurement of 'stability constants' of complexes is important in biology and in industry. Thus, it is found that certain drugs (organic compounds) are bactericidal only in the presence of certain metal ions. In order to know whether it is the metal complex that is actually involved in the reaction one must determine what fraction of the drug is actually bound to the metal ion at the acidity of the biological system. Certain antitubercular drugs, for example, are believed to penetrate the fatty layer around the bacteria as a non-electrolyte metal-complex. To test this theory at a given acidity one must determine the extent of the reaction

metal ion + drug ⟶ metal complex.

Another example is from the nuclear energy field. After the uranium fuel has reacted for a suitable time in the pile it is removed, dissolved in nitric acid, and the fission products separated. One of these fission products is ruthenium, which dissolves to form various complex compounds. In the early stages of the development of the process ruthenium caused difficulty because its complexes slowly changed into others with different physical and chemical properties, and the ruthenium tended to be deposited in more than one stage of the separation process – a headache for the chemist and chemical engineers on the plant. Not only did we not know in 1940 the rate at which these complexes reacted and the extent to which these reactions proceeded, but even the nature of the ruthenium nitrate complexes was practically unknown. When we recall that there are many elements besides ruthenium about which we know little of the nitric acid chemistry one sees how much work is still to be done.

CONCLUSION

Enough has been said to show that modern inorganic chemistry is an integrated subject with much more unifying theory than was the case in years gone by. It also offers an exacting and rewarding field for research workers. In later issues some of the avenues for development will be discussed in more detail. There is at the present time an appalling shortage of research workers in inorganic chemistry, but it is certain that the presentation of the subject at schools, technical colleges, and universities, in the true light, will do much to remedy this situation.

BLOOD

M. C. H. DODGSON

IF we imagine the animal body to be a country, or a single economically integrated unit, the blood stream would be its sole transport system for the conveyance of raw materials and waste products to and from the innumerable factories scattered about the land. Oxygen, a commodity that is required by every 'factory', or tissue, is carried round the body in 'barges' – the red blood cells. When these are empty, a universally occurring waste product, carbon dioxide, is conveyed back in them for disposal. Except for these two gases, all other chemical substances – digested foodstuffs, internal secretions, acids and alkalis, drugs, chemical protectors against infection (antibodies), waste products – are carried round the body dissolved in the waters of the canals, to be selected or rejected by each separate tissue according to its needs. To go further into the nature and destination of these constituents of the *plasma*, as the continuous phase of the blood is called, would be to attempt to epitomize the greater part of modern biochemistry and physiology. Haematology, on the other hand is concerned more with the production, upkeep, and repair of the blood stream itself.

In a system of land canals, although the barges are valuable vehicles, the water itself is as a rule in unlimited supply and its conservation presents no special difficulties. The blood stream however is pumped around the canals at a pressure considerably above atmospheric (with each beat of the heart, the pressure of blood in the larger arteries rises to approximately 120 mm. of mercury), and consequently every drop is precious. When blood escapes from the vessels, in a wound for instance, unless a major artery has been severed, relatively little blood escapes and the leakage soon ceases. One of the means of ensuring this spontaneous arrest of haemorrhage, or *haemostasis*, is

the fact that the injured blood vessels contract, greatly reducing for a short time the flow of blood into the injured area. Contraction of vessels, or vasoconstriction, is however soon followed by their relaxation, which increases the blood flow to the damaged area, constituting the first part of the process of tissue repair. Thus if transient vasoconstriction were all that happened, we should be in danger of bleeding to death after cutting ourselves whilst shaving, or following dental extractions and even the most minor surgical operations. Such is indeed the state of certain unfortunate people known as haemophiliacs. That a cut does not normally continue to bleed indefinitely is a consequence of the solidification of blood that takes place within a few minutes of its being shed. As a result, the mouths of severed vessels become sealed and the opposite edges of small wounds are held together, thus precluding further haemorrhage. In haemophilia, the blood, through an inborn chemical defect, cannot clot adequately and in severe cases may remain fluid indefinitely.

BLOOD CLOTTING

Unlike the setting of a gelatine solution on cooling and its subsequent liquefaction when warmed, the clotting of blood is not a reversible physical change, but is the irreversible result of a minor chemical alteration in one of the proteins that is dissolved in the plasma. This chemical modification of *fibrinogen*, with the formation of a new substance, *fibrin*, does not take place instantly in any circumstances but occurs at the end of a chain of chemical reactions, each successive reaction being fired off by some product of a previous one. The initiation of the series of reactions is very readily caused, in a number of different ways, and must happen from time to time within the circulation of a normal healthy person. Nevertheless, since the blood is in constant motion within the body, it must be a rare event for any substance capable of setting off the clotting of blood to be able to accumulate in sufficiently high concentration for the process of clotting to continue. Apart from this dilution effect, there is a second protective mechanism against the occurrence of blood clotting within the vessels, or *thrombosis*; when it has formed

within blood vessels in small quantities, fibrin may be redissolved by a process of *fibrinolysis*, involving a series of chemical reactions which are at present little understood. Thrombosis of veins or arteries does of course occur at times, when a part of the blood stream is temporarily sluggish, and clotting substances that may have formed fail for that reason to become so diluted as to be harmless. Thus clots may form in the veins of the legs when the vessels into which they flow are partially obstructed in the pelvis during pregnancy or when the circulation of the legs is sluggish during the immobilization in bed that follows a surgical operation. Similarly in persons with disease of the coronary arteries of the heart, coronary thrombosis is most likely to occur during sleep, when the activity of the circulation is minimal.

After the preliminary reactions, the change that is responsible for the conversion of liquid to solid blood is the alteration of fibrinogen to fibrin. Fibrinogen, in this respect resembling gelatin, is one of those proteins in which it has been possible to show, from a consideration of the nature of its X-ray diffraction pattern and electron microscopic appearances in relation to molecular weight and to the nature of the chemically reacting 'side chains', that the molecule consists of a long polypeptide backbone and is itself already a miniature fibre. Polypeptides are compounds formed by the combination of two or more amino-acids, which, bearing an acidic group at one end and a basic radical at the other, are readily linked in this manner. Polypeptides constitute the bricks of which the protein molecule is built. It is thought that in many instances these *macro-molecules* have a helical, or spiral, configuration, like a coiled spring. It has been said that a protein is essentially a pattern of charges associated with a closely packed forest of side-chains, the polypeptide 'backbone' being an infinitely varied device for building up distributions of charges and presenting them to the environment.

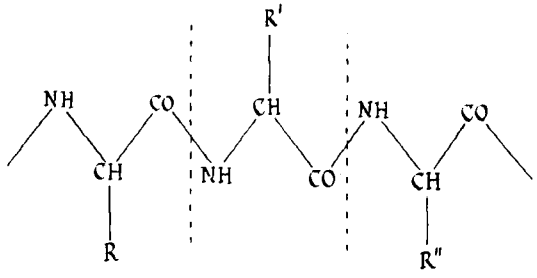


Fig. 1. Part of a polypeptide chain comprising the amino acids.

Fibrinogen molecules exist singly in solution in the plasma. The exact nature of the change they undergo during the clotting of blood is not fully understood. Fibrin, the insoluble 'skeleton' of a blood clot, is a meshwork of fibres, formed by an infinitely ramifying aggregation of modified fibrinogen molecules. It is the nature of this modification of the fibrinogen molecule that has given rise to difficulties. A true solution of fibrin may be prepared by dissolving a fibrin clot in urea. When this is done it is found that the molecular weight of the fibrin is slightly less than that of the fibrinogen from which the clot was originally prepared. From this and other evidence it has been suggested that the polypeptide backbone of the fibrin molecule is slightly shorter than that of fibrinogen. In other words, the culmination of the chain of chemical reactions which precede clotting is probably the 'biting off' of the terminal group of one end of the fibrinogen backbone, leaving a more reactive molecule which, by interacting with similarly prepared molecules, rapidly forms a *polymerization* compound. The nature of such compounds will now be described.

A polymer is a substance consisting of molecules which are, at least approximately, multiples of units of low molecular weight. An almost universal example of polymerization is the production of polystyrene compounds. The parent substance, styrene, comparable to 'activated' fibrin, has a relatively simple molecule, consisting of a benzene ring to which is attached a short chain of two carbon atoms, united by an unsaturated valency bond ($C_6H_5.CH=CH_2$). Styrene itself is a colourless aromatic fluid; by appropriate chemical treatment, however, infinitely long chains of these molecules may be formed by condensation. The resulting polystyrenes are solid substances

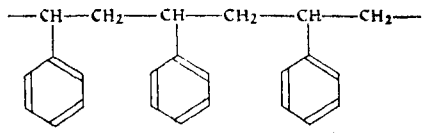


Fig. 2. Three units of a polystyrene chain.

from which are manufactured many commodities, including toys, wall tiles, refrigerator parts, radio cabinets, tooth brushes, and even optical lenses.

Now that we know something of the final precipitation of fibrin, it will be as well to return once more to the beginning of the process of blood clotting. When blood is withdrawn with a hollow needle from a vein and allowed to run into a dry test-tube, nothing appears to be happening for about four and a half minutes, until quite suddenly the whole sample solidifies to form a soft, dark red jelly. A most complex series of chemical events is set in train, however, immediately the blood has been withdrawn from the circulation. In its inevitability and abruptness, the clotting of a sample of blood may be compared to the striking of a clock, once the preliminary whirrings of the sounding mechanism have begun.

In addition to the cells of the blood stream, there are also suspended in the plasma large numbers of tiny bodies, known as *platelets*, which are of about the same size as bacteria. These bodies are sticky; as soon as blood comes in contact with a surface other than that of the blood vessel lining – the inside of a test-tube, for instance – clumps of platelets adhere to the foreign surface and liberate one of a class of chemical substances, collectively known as *thromboplastin*. Platelet-formed thromboplastin is a compound of fatty substances with protein, known as a *lipoprotein*. In normal individuals the liberation of thromboplastin occurs almost instantaneously in shed blood. In haemophiliacs, however, the lack of one of the plasma protein fractions which normally acts as an accelerator to the production of thromboplastin from platelets and which, for want of a better name, is known as *anti-haemophilic globulin*, results

in a serious delay in the initiation of clotting. Since the various substances involved in these important preliminary reactions are so labile, relatively little is known of the chemical mechanisms that are involved. Other 'thromboplastins', apart from that which is liberated from platelets, are derived from damaged tissues of all kinds, particularly brain and muscle. These initiators of clotting act rather less promptly than platelet-thromboplastin and are phospholipids (of which cephalin and lecithin are examples), not lipoproteins. They are of importance in helping to ensure the clotting of blood in a wound.

Also among the circulating proteins is a carbohydrate-containing component, *prothrombin*, which plays a central part in the next stage. In the presence of ionized calcium, of a further globulin component of the plasma, and probably of other factors, thromboplastin converts prothrombin, by what must be a relatively minor change in its molecule, to *thrombin*. This last substance now reacts directly with fibrinogen, changing it to fibrin. The clotting of blood is therefore brought about in three stages, by a chain of chemical reactions which take a little over four minutes to complete and which, except for the last, require the aid of catalysts. Summarizing, these three stages are: (*a*) the release of thromboplastin from platelets or from damaged tissue, (*b*) the conversion, by thromboplastin, of prothrombin to thrombin, (*c*) the conversion, by thrombin, of fibrinogen to fibrin.

A few practical applications of this knowledge may be mentioned. In haemophilia, where the defect in the clotting mechanism is due to an inability to produce sufficient quantities of thromboplastin, symptomatic of an absence of small quantities of a specific 'anti-haemophilic' globulin in the blood, this last factor, which is unstable, can be given in sufficient amounts to a haemophiliac who is bleeding to correct the clotting defect and stop the haemorrhage, simply by giving the patient a transfusion of fresh normal blood. Blood stored in a refrigerator for any time is not so effective, since the required factor is so labile.

The storage of blood for blood transfusion can only be achieved if clotting is inhibited. This is done most effectively by

removing the ionized calcium from the blood, thereby totally arresting the process at stage (*a*), since the conversion of prothrombin to thrombin does not occur in the absence of calcium ions. Therefore blood donors are bled into large bottles containing small quantities of a solution of sodium citrate; in this way the calcium ions are captured, with the formation of unionized calcium citrate.

It is sometimes necessary to reduce the coagulability of a patient's blood stream, following coronary thrombosis for instance, or in cases of inflammation or thrombosis of veins. This is done quite simply nowadays by giving one of a number of drugs by mouth, which impair the clotting mechanism by reducing the rate of production of prothrombin by the liver. These *anticoagulant* drugs have replaced a number of other anticoagulant substances such as heparin which had to be given by injection directly into a vein to produce the desired effect.

THE ERYTHROCYTE

The red blood 'cells', red blood corpuscles, or erythrocytes, which are the oxygen carriers of the body, are not properly speaking cells at all, since they have lost their nuclei and are simply compact and exceptionally dense fragments of protoplasm. In shape they are something like flying saucers, but differ from the usual conception of these vehicles in being biconcave. Although they are described in text-books as being symmetrical in section, they very often appear to be more concave on one side than on the other, i.e., more truly saucer-shaped than is generally supposed. Unlike space ships, however, they are so numerous and closely packed as to jostle one another constantly on their journeys through the blood vessels of the body. Their average diameter, a little over 0.007 mm., is appreciably greater than the calibre of many of the smallest capillary vessels through which they must pass and they are therefore subjected to further mechanical stresses through bending or folding in these vessels. Having no nuclei, the erythrocytes are without nucleoprotein, whose molecules act as 'templates' for the synthesis of the proteins with which living protoplasm is constantly being renewed.

The life of an erythrocyte is therefore limited: it is perhaps surprising that they are able to withstand the shocks and buffetings of the circulation for as long as four months, which is their average life span. A single erythrocyte may circulate round the body as many as a hundred times in one day.

At one time it was supposed that the red cell was a minute drop of fluid surrounded by a membrane of fatty substances and that it owed its peculiar shape to the combined effects of surface tension and of some sort of internal scaffolding, around which the cell sap circulated. It is now known that the protoplasm of the erythrocyte, containing only 65 per cent of water, is considerably denser than most forms of living matter and that it exists in the form of a loose aggregate of protein crystals, on the surface of which there is a monomolecular layer of lipids, mainly lecithin. Although it is freely fluid, as can be observed with the microscope, the interior of the normal red cell is probably not a true solution, since it is likely that the forces of mutual attraction between the protein molecules, these in a sense resembling magnets in having opposite electrical charges at either end, are responsible for maintaining the shape of the corpuscle. In certain circumstances, however, the interior of the red cell may become a true solution. When this occurs, the shape

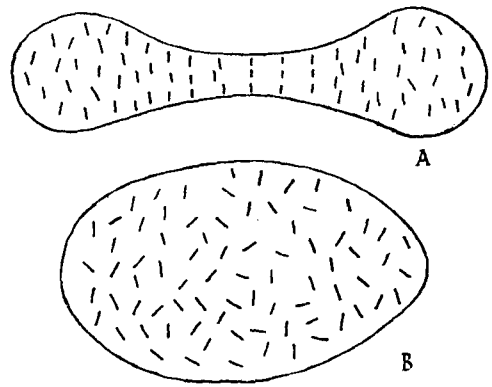


Fig. 3. Erythrocyte (A) and Spherocyte (B).

B

of the cell is determined solely by surface tension and it tends to become spherical in outline. Once this has occurred, the *erythocyte* (A) is known as a *spherocyte* (B); now being a drop of protein solution surrounded by a delicate semipermeable membrane, it will continue to swell and eventually bursts if it is placed in solutions of progressively lower osmotic pressure.

From the science of crystallography we know that the characteristic shape of a crystalline solid is determined by the spatial arrangement of the atoms in its molecule. The angles at which the faces of the crystals are set are determined ultimately by the structure of the molecules of which the crystal is composed. It has been recognized recently that many proteins are crystalline in nature. X-ray analysis and other methods of investigation have shown, for instance, that the molecule of ***haemoglobin***, the red protein-containing pigment which comprises some 90 per cent of the total solids of the erythrocyte, is made up of four superimposed layers or discs, composed of folded polypeptide chains, with two very much smaller, flat, rectangular 'haem' groups applied to their surfaces.

In the centre of each of these haem rectangles there is an iron

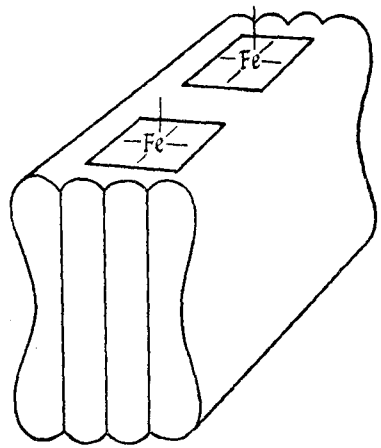


Fig. 4. Probable structure of a molecule of horse haemoglobin (after Granick and Gilder, 1947).

atom, held by all save one of its six co-ordination links, which is thus left free to combine loosely with an oxygen molecule. This is the structural basis of the oxygen-carrying property of blood.

Although it is hardly true to say that the erythrocyte is an enormously magnified haemoglobin molecule, it is at least possible that the fact of its being a biconcave disc is an expression of the structure of its molecule. Throughout the animal kingdom there are many different types of haemoglobin, differing principally in the composition of the protein component of the pigment; there is also a certain degree of variety in the shape of the red cell of different species; the camel, for instance, has oval erythrocytes. To what extent these morphological peculiarities are related to individual peculiarities in the structure of the protein part of the haemoglobin, is at present conjectural.

The differences in the physicochemical properties of the various types of haemoglobin that have been recognized depend for the most part on variations in the mode of folding of the polypeptide chains of the protein part of the molecule, the 'globin'. All varieties of haemoglobin have a molecular weight of approximately 66,000, so it is unlikely that there are any wide differences in the amino acid components of these chains. In man, the type of haemoglobin that is synthesized by the foetus is particularly well suited to conditions within the uterus, where oxygen is brought to it by the maternal blood stream at considerably lower tensions than will obtain in the infant's lungs as soon as it has been born. Human *foetal haemoglobin* has a greater affinity for oxygen than the adult type and it releases carbon dioxide more readily. The change from one form of haemoglobin synthesis to the other occurs immediately after birth. Red cells containing the foetal type of pigment are rapidly destroyed and the child may become jaundiced on or about the third day of life as a result of the accumulation in the blood stream of bile pigments derived from the excess haem. At the same time the new mode of synthesis may be inefficient at first and slow to get under way, particularly in prematurely born babies, who are for this reason liable to become anaemic during the early months of life.

It has only been during the last few years that the existence of varieties of human haemoglobin, apart from the normally occurring foetal and adult types, has been recognized. Now, in addition to these two, five other distinct types of haemoglobin are known; they are present in the blood stream in only a minority of the population. The significance and mode of inheritance of these anomalies of protein synthesis is being studied by a number of research workers. A phenomenon that has been recognized for many years is the *sickling* of the erythrocytes of certain individuals, principally Negroes; cells that were previously circular, and to all appearances indistinguishable from normal erythrocytes, become narrow crescents in circumstances where the oxygen in the fluid surrounding them is reduced. It has been shown that the haemoglobin of persons with this sickle-cell trait is of a peculiar type which is, in its reduced state, i.e., when the iron of the haem complex is not attached to oxygen molecules, one hundred times less soluble than when it is holding oxygen. Reduced normal adult haemoglobin, on the other hand, is about half as soluble as the corresponding oxygenated pigment. Thus the sickling of red cells is thought to indicate the crystallization of this abnormally insoluble haemoglobin. Fortunately the phenomenon rarely occurs in the bloodstream *in vivo*, as when it does it leads to the destruction of the abnormal erythrocytes, with the production of a severe anaemia. Only persons who are heterozygous for the gene, i.e., who have inherited the tendency to form sickle cells from both parents, are liable to develop anaemia as a result of sickling *in vivo*. An interesting side issue to the story of sickling is that the possessors of this type of haemoglobin are relatively immune to malaria, the parasite presumably finding conditions inside these particular red cells to be relatively inhospitable.

THE BONE MARROW

The cells of the blood and the platelets occupy just less than one half of the total blood volume, i.e., they would altogether fill between 5 and 6 pints. The number of red cells in the circulation of a fully grown adult is of the order of 20,000,000,000,000, plus

30,000,000,000 leucocytes and about 600,000,000,000 platelets. These seemingly astronomical figures may be surprising till one remembers the relatively minute size of the cells of which the solid parts of the body are made. Although they are by no means accurate estimates, they serve to give a vivid illustration of the immense difference there may be between the size of an organism and that of its parts. Built on this sort of scale, a skyscraper constructed of 3-in. bricks, each brick representing a cell, would tower upwards to a height of three to four miles. It is more important, however, to realize another implication of these huge figures, namely that each cell is formed as a result of a fairly complicated process of protein synthesis and that the whole process is gone through in millions upon millions of separate foci. In order to make up for losses occasioned by normal erythrocyte ageing, wear and tear, and small haemorrhages, some 9,000,000,000 erythrocytes are produced every single hour. The relatively enormous 'factory' that is required to house this industry is scattered throughout the body in the medullary cavities of the bones, in which there is room for as much as a tenfold expansion should the needs of the body, e.g., following a large haemorrhage, demand it.

The maintenance of a concern of this size in full production depends on the constant provision of the necessary raw materials required for the formation of blood cells. Indeed so great is the activity of the bone marrow that even in normal health there is little reserve of resources. A constant drain on the stores of iron, such as occurs after repeated small haemorrhages (e.g., even the normal menstrual flow during the child-bearing period in women) may result in a deficiency of this essential component of the red blood cell and cause anaemia.

At the end of their life span, erythrocytes break up, mainly in the spleen, and their haemoglobin is liberated. It is immediately split into haem and the protein, globin; the latter becomes part of the protein stores of the body, while haem is further degraded in the liver into a variety of non-iron-containing compounds which are excreted in the bile, the iron, however, being jealously retained by the body to be used by the bone marrow in the

resynthesis of haemoglobin. In this latter process, the components of the pigment are synthesized separately, the polypeptide discs, which have been laid down in the immature red cells, capturing the iron-containing haem groups by virtue of the electric charges on the surfaces of the protein molecules. This is part of the final maturation of the red cells and is extremely sensitive even to minor deficiencies in the amounts of available iron. Should there be a lack of iron for any reason, the supply of erythrocytes is numerically unaffected, but these cells are turned out of the bone marrow paler than normal, i.e., possessing less than their proper quota of haemoglobin. Measurement of the colour of the blood is used in anaemic patients as a means of estimating the concentration of haemoglobin present, and hence of the degree of iron deficiency.

How, in health, the bone marrow succeeds in turning out no more and no less than the correct number of red cells into the blood stream to maintain a constant concentration of approximately five millions per cubic millimetre, is by no means adequately explained. It is clear, however, that the successful multiplication of cells must depend on there being an adequate supply of protein, which is brought to them via the plasma and whose source is ultimately from the diet (following the digestion of proteins to their constituent amino-acids and their resynthesis to proteins of human pattern in the liver). Deficiency of protein in chronic malnutrition, particularly in children and in pregnant women in the tropics, is of common occurrence and declares itself, amongst its other manifestations, as a severe anaemia. Although, no doubt, the protein lack may be such that not enough is present for the marrow to be able to build from it an adequate supply of cells, the responsible defect appears to be the lack of a particular component of the vitamin B complex, B_{12}, which is present in meat and grain and contains in its molecule the element cobalt. It is inevitable that this vitamin will be missing when the supply of protein in general is low.

The result of lack of vitamin B_{12} from any cause, of which dietary protein deficiency is only one, is what is known as *megaloblastic anaemia*. This is only a way of saying that when

a sample of bone marrow is removed with a syringe and examined microscopically, it is found that there are many more than the usual numbers of immature cells (they are not yet 'red' cells because they contain no haemoglobin), that these cells differ in appearance from the normal healthy erythrocyte precursors, and that the relatively few that succeed in maturing as red cells tend to become rather larger than they are in normal circumstances. It has recently been discovered that there may also be a reversion to the foetal type of protein synthesis within the immature red cells, and that of the few erythrocytes that appear in the circulation, a proportion contain the foetal type of haemoglobin. The marrow may be literally stuffed with immature non-haemoglobinized cells, while no more than a thin trickle of fully mature erythrocytes ever reaches the circulation. It is as though in a factory producing cars on a mass scale, a strike of the workers concerned with one of the finishing processes were countered by an obstinate continuation of a previously high rate of production up to this essential final process, as a result of which the factory soon became cluttered with unfinished and useless vehicles. In some cases within a few hours of giving a dose of vitamin B_{12} to a patient with a megaloblastic anaemia, the 'finishing' bottleneck is overcome and the orderly normal mode of production of erythrocytes is resumed.

PERNICIOUS ANAEMIA

Until 1926, this severe megaloblastic anaemia in middle-aged and elderly persons could not be treated and was inevitably fatal. Following the termination of the 1914–18 war, Whipple and his colleagues had begun to work to a research programme in which dogs were made anaemic by bleeding, after which the rate of new blood formation could be studied. By feeding the animals in different ways, it was soon realized that of all foodstuffs tried, liver produced the most rapid and satisfactory recovery from the anaemia. In 1926, Minot and Murphy, appreciating the value of this work, began to feed liver to patients with pernicious anaemia and obtained complete recoveries, although to maintain their erythrocyte production the patients had to con-

tinue the diet for the rest of their lives. Initially it was given raw and in large quantities.

The appearance of symptoms of anaemia was known to be preceded commonly by a period of twenty or more years, during which time a variety of gastric disorders had been experienced. It was also known that in patients with pernicious anaemia the normal highly acid gastric juice was not produced, there being in every case a severe degree of atrophy of the epithelial cells lining the stomach. It thus seemed probable that the deficiency lay not so much in an inadequacy of diet as in the failure of the stomach to complete some digestive change that was necessary in order to render some foodstuff capable of maintaining normal blood cell production.

Further research showed that, given adequate supplies of protein and iron, all that is required in the diet to make normal haemopoiesis possible is the constant intake of minute amounts of a substance which, before it had been chemically isolated, was originally known as 'extrinsic factor', but which is now known to be vitamin B_{12}. When, as in pernicious anaemia, the glands of the stomach have ceased to produce gastric juices, the presence of this substance in the diet does not give rise to a normal production of red cells. This is because the normal stomach elaborates a ferment, 'intrinsic factor', which alters the molecule of B_{12} chemically in such a way as to render it capable of stimulating the bone marrow to produce a normal supply of cells. Unmodified B_{12} has no such action. This altered B_{12}, known before its isolation as the 'haemopoietic principle', is stored in the liver, whence it is liberated into the blood stream for utilization by the bone marrow. The curative effect of liver-feeding in patients with pernicious anaemia is thus explained.

We are now in a position to understand the mode of production of megaloblastic anaemia – the blood changes being in all instances essentially similar – in a wide variety of different circumstances. In pernicious anaemia the slowly developing atrophy of the mucous membrane of the stomach results finally in a failure to secrete intrinsic factor, leaving the adequate supplies of B_{12} unconverted to a state in which they can be made

use of by the bone marrow. Secondly, following surgical removal of the whole of the stomach, certain patients may be left in virtually the same state as are sufferers from pernicious anaemia and may develop, years after the operation, the same blood disorder. A third group of cases is somewhat unexpected. It is well known that the intestine contains millions upon millions of bacteria. Bacteria, like the cells of the bone marrow and probably all other dividing cells of the body, require supplies of vitamin B_{12} in order to multiply. There is thus a competition for this vitamin between the bacteria of the gut and the cells of the bone marrow. The bacteria, of course, have first call on the supplies and only that which is left over can be absorbed by the gut and stored in the liver. The actual numbers of bacteria in the intestine may therefore play a decisive part in precipitating a vitamin B_{12} deficiency, particularly when the supply from the diet is on the margins of inadequacy.

There are in Britain, in addition to those that have already been mentioned, many other rare forms of megaloblastic anaemia, most of them associated with chronic intestinal disorders. It is, however, in the tropics, where anaemia may be of almost universal incidence and may contribute materially to the backward condition of a nation, that the importance of the nutritional megaloblastic anaemias is vividly demonstrated. Other potent sources of anaemia, hookworm disease, chronic malaria, and in some areas, schistosomiasis, can be eliminated from a locality by taking appropriate and energetic hygienic measures. The problem of the nutritional megaloblastic anaemias of pregnancy and in childhood, a direct consequence of protein deprivation, has its roots in the economic structure and in the life of the country. Although individual cases can readily be treated, the disease as a whole, with all its consequences of retardation and ill health, will prove difficult to eradicate.

LEUKAEMIA

In contrast to the thoroughness with which the chemical control of red cell formation has been investigated, there is at present little that can be said of the control of the production of

leucocytes by the bone marrow and in the lymph nodes. So wide are the variations in the concentrations of these cells in the blood of different individuals, and sometimes in that of a single subject from hour to hour, that the balance between supply and demand does not appear to be so finely adjusted as it evidently is in the case of the production of erythrocytes. The rate of turnover of the leucocytes is considerably more rapid than that of the erythrocytes: those formed in the bone marrow live for about three weeks from their liberation into the blood stream, while the life of a lymphocyte (formed in the lymph nodes) may be measured in hours.

The principal disorder of leucocyte formation, leukaemia in its various forms, has been much studied but little understood. In certain animals this disease, shown by the presence of excessive numbers of fully formed or immature leucocytes in the peripheral blood, may be the response of the animal to infection with a virus. In such circumstances the disease may be transmissible experimentally from one animal to another via cell-free filtrates of leukaemic blood. Although there is reluctance to accept the possibility of human cases being of an infective origin, it must be pointed out that attempts to transmit the disease from man to man, rather than from man to animal, to exclude the presence of an infective agent, have not been carried out, nor are they likely to be.

Despite our ignorance of the cause of this disease, we do know a little about the circumstances in which it may develop. Like that of carcinoma of the lung, the incidence of leukaemia has increased in recent years, being responsible for a little over 400 deaths in England and Wales in 1920, as against 2,200 in 1952. The increase still continues. Certain classes of workers, e.g., those exposed to X-rays or radioactive material, or who handle benzene and its derivatives, are significantly more prone to develop this form of blood disease than are members of the population as a whole. With regard to the effects of radioactivity, there is known to have been an incidence of leukaemia among the survivors at Hiroshima and Nagasaki, of approximately 50 times that in the Japanese population at large.

It is therefore probable that the recent increase in the disease, if it is not due to the presence of a virus of low infectivity, represents the response of the bone marrow to other environmental factors. Apart from the increased prevalence of radiation in all forms, including X-rays, during the last thirty years, a feature of medical practice during this period has been the introduction of new drugs, many of which, containing the benzene ring, are known to have an effect on the bone marrow. Occasional cases of leukaemia immediately following the use of sulphonamides have been reported; it is thus possible that the use of these and other drugs has contributed to the increased incidence of the disease.

The leukaemic process in man is generally considered to be neoplastic; that is to say that the disorderly proliferation of cells in the bone marrow which is responsible for the excess of leucocytes in the blood and for the accompanying anaemia, caused apparently by the crowding out of normal erythrocyte-forming tissue, is regarded as being a form of tumour. Certain reservations should be made, however, since persons with chronic leukaemia, particularly lymphocytic, may live in good health for years despite their greatly increased white cell counts. Wilkinson and his colleagues champion the view that the category 'leukamia' includes two distinct diseases. In the acute form, the predominant cell is characterized by its immaturity and by its atypical features, which are reminiscent of those seen in cancer cells. There is little dispute as to the neoplastic nature of cases of this kind, which if left untreated are rapidly fatal. In the chronic form, however, the cells are mature ones and are identical with those occurring, although in lesser numbers, in the normal bone marrow. It is therefore possible that in cases such as these the initial disturbance is one of the storage of leucocytes, rather than an excessive production. However this may be, the only effective treatment of the leukaemias at the present time is by the use of drugs which exert a specific inhibitory effect on cellular division. These at least offer the hope of arresting the disease process.

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SPEECH ENGINEERING

D. J. BRUCE

MAN'S use of language has interested philosophers for many centuries, but only in recent years has it become the subject of experimental investigation. The experimental approach has resulted in a change of emphasis. Although features such as meaning and concept formation are still of central importance, they are no longer the starting point for analysis. In studies of speech perception and language engineering the tendency today is to begin at the beginning with phonation and audition, and work up to the ultimate implications of verbal behaviour. Indeed, it is often convenient to forget about the higher aspects of language and concentrate on speech as a perceptual-motor skill.

There is considerable justification for this limited approach. Within its terms of reference a whole host of practical problems can be studied – hearing defects, speech difficulties, communication systems, and so on – all as variations on one general theme. Consistencies rather than the differences of language then stand out, and the systematic account of language that is emerging from this work should result in a more direct attack upon semantic problems.

Outside the laboratory, each of us tends to take his powers of speech for granted, but it has actually taken us something like fifteen years to achieve mastery of our mother tongue. It is instructive to speculate on how much poorer off we should be had our verbal capacity been more limited. Suppose we could make only two sounds – let's call them *0* and *1*. It would be possible to create a fully descriptive language on this basis, the Morse Code is such a language, but we would have to increase our talking time alarmingly to make the simplest statements. 'Words' would be very lengthy, not only because they would

consist entirely of *0*s and *1*s, but because we would have to introduce pauses *within* words to avoid confusion of sound sequences. Thus the word BIB might become *01*-pause-*1101*-pause-*01*.

This sort of language would at least allow us to cope with the business of representing the external world, however laboriously, but what if our limitation were that we could not combine the sounds we can make? Miller[1] gives an amusing example of such a language. 'Ah' might stand for a greeting, 'ee' for a request for help, but by the time we had used less than a hundred different sounds we would have covered the range it is possible for man to make and distinguish, and correspondingly drained the pool of potential symbolization. To be adequate for communication, speech must represent not a hundred but many thousands of different things, and must achieve this representation with economy of time and effort. When we examine the stream of speech we can see how this has been done.

The sounds we make are combined in various patterns, and, fortunately, we can make sufficient sounds to provide a very great number of patterns without the difficulties attendant on the imaginary two-sound language. We call these patterns words. The average college student has a vocabulary approaching 156,000 words, which gives considerable opportunities for representation compared to the isolated sound language. But this sort of combination is only the beginning of the patterning shown in our speech. Words themselves are combined and arranged in different orders to provide refinement in description, and to achieve the representation of more complex situations. This not only increases the scope of language, but it is economical as well. To take the simplest example, our vocabulary would become top-heavy if we used a different word for each condition of a particular object. Instead of this, in the English language we use qualifying words – *red* sky, *cloudy* sky, *clear* sky, *stormy* sky – which can also be employed to describe similar conditions of other objects. So we can trace a scale of patterning in our language running up from the simplest linkage of sounds in what are known as morphemes, through words, phrases, sentences, and beyond. At each stage the patterning

becomes more intricate and its quality of representation more discriminating.

Our language is a powerful and complicated tool. Obviously for us to use it confidently, as we do, we must have certain rules of operation. It is here that the problems come in. We are all aware of a few rules that govern the structure of our speech – points of grammar and syntax, social permissibility and so on – but if we study them they are revealed as surprisingly inadequate as an explanation of the general accuracy and fluency of speech. If I say to you now 'I hope you are finding this article . . .', and ask you to complete the sentence, you have no difficulty. You would probably supply 'interesting', and you would be right in the sense that that was what I should have said had I finished it. You might say something like 'enjoyable' or 'stimulating', in which case you would be very nearly right. It is highly unlikely that you would complete with 'foolish', and practically certain that you would not say 'a nightmare'. Why?

To answer this question it is necessary to introduce the idea of *constraint*. By this is meant a reduction in freedom of choice in speaking or assimilating what is spoken. It comes about in a number of different ways. Thinking about the above example, we can recognize the effect of at least three different forms of constraint. The most obvious is syntactical. In the course of learning to speak we have acquired a stock of permissible patterns for the parts of speech that are recognized in our language. We know that we can use the pattern *noun verb noun*, for instance, but that *noun noun verb* is rarely permissible. Thus 'Cat eats mouse' is a form with which we are abundantly familiar (and expect to encounter), but 'Cat mouse eats' strikes us as strange and not quite right because its use is so restricted. So, in the present example, your knowledge of syntax leads you to expect an adjective as the most likely completion of the pattern 'I hope you are finding this article . . .'. This constraint is not sufficient in itself, however, to explain the readiness of your response. There are many adjectives from which you could choose, but your choice is still further limited by another constraint. This has been called *verbal context*.

At this stage of the argument we come very close to the field of semantics, though it is not absolutely necessary to introduce questions of meaning, as I shall show later. For the present it is sufficient to note that, by virtue of what has gone before, your choice of adjective will be relevant. That is, it will be the sort of word that can be associated with writing, articles, and so on. The number of alternatives for your response is now considerably reduced, but still too extensive for us to say the process is complete. On the basis of constraint so far, you could as well say 'printed' as 'interesting'. There must be yet another set of constraining factors which reduce choice to a more manageable state.

It is convenient to label these remaining constraints as *situational context*, though the term embraces so much as to be of doubtful use. It includes judgments on your part which go outside the immediate verbal context and which enable you, for instance, to anticipate a word that will have reference to the impression made by my writing rather than its physical qualities. Moreover, you know from your experience of the relations between writer and reader that the type of impression I am hoping for is a favourable one. Add to this your knowledge of the restraint employed in scientific discussions, which eliminates alternatives such as 'breath-taking' or 'ecstatic', and you are left with a practicable choice situation. It is still far from simple, and you have to be a bit arbitrary about which of its dimensions you pick. Am I referring to intellectual or emotional appeal? You decide the former, and your final selection along this dimension is probably governed by frequency of usage of the remaining alternatives such as 'absorbing', 'intriguing', or 'interesting'.

The startling thing is that this whole long drawn out process of selection occupies next to no time and you are barely conscious of it. I make my unfinished remark and its completion comes to you seemingly ready made. Yet you perform this verbal *tour de force* not just when requested, but constantly, whenever you listen to someone speaking. The perception of speech is characterized by a continual running anticipation on the part

of the listener of what is still to be said – an anticipation that shifts and changes according to what has already been registered. Of course this anticipation does not usually involve spoken completions, though you have probably met individuals who cannot resist the temptation to finish your statements for you. The process is not really an expressive one, but a means of dealing as efficiently as possible with the formidable stream of stimuli represented by a speaker's output. If we were denied this active, predictive approach and reduced to registering speech word by word, we should soon be lost. It has enormous advantages, and the fact that it also leads to errors of apprehension is usually of small moment for most people.

One of the most important problems we have to solve, then, for the understanding of speech perception is how these constraints operate to enable the listener to make his continuously modifiable predictions. Stating the problem in another way, we have to assess how he employs contextual cues, and what forms those cues take. I shall describe some experiments which attack the problem, but I realize that they only touch its fringes.

I have already mentioned one of the many difficulties that beset the worker in this field, the rapidity and elusiveness of the process he is trying to study. He has in some way to slow it down, and the easiest way of doing this is to make the listener's job harder. So the subject of these experiments is usually asked to listen to speech under noisy conditions. This may not only produce the desired effect, but the results are often pertinent to the improvement of communication in factories, military and other situations where noise is a prominent and inevitable factor. The noise employed in the experiments is not any noise, however. Remember, it will be used for the purpose of masking speech and it is important that no part of the spoken material should be favoured in its intelligibility over any other part. This means that not only must the noise be continuous as long as the test material is being spoken, but that it should be an adequate masking for all the frequencies involved in the speech spectrum. In practice, this is effected by using white or random noise which is made up of all the audible frequencies in equal amounts. It sounds

rather like the hissing of surf, though of course it does not wax and wane. There are several ways in which such noise can be produced. One of the simplest is to amplify the results of placing an ordinary magnet around a live thyratron. The noise and the verbal material to which the subject must listen – usually called the signal – are mixed and conveyed to the subject's ears through headphones. He is asked to record, either orally or by writing, what he hears.

Some idea of the advantage accruing from the use of contextual cues can be gained from an experiment of Miller, Heise, and Lichten's.[2] In this test subjects listened to the same set of words on two occasions, but with this difference. On one occasion the words were given in isolation, on the other they were embedded in sentences that provided the type of verbal and syntactical context I have described, thus fostering an active anticipatory approach. You can see from the graph given below how the presence of contextual cues improved the intelligibility of the words. The graph's baseline refers to the customary way of expressing the relative loudness of signal and noise, using decibels as the unit. Positive values indicate that the signal was louder than the noise, negative values the reverse. The results of this experiment might be summed up by saying that a listener can hear under suburban street conditions (signal-to-noise ratio +12 db.) 30 per cent more of what is said to him if he is allowed to make use of context than if he is denied its aid.

A study of this type is valuable for demonstrating the general importance of context to speech perception, but it is not in itself very discriminating. You can see that the method employed allows no way of estimating the relative contributions of syntactical and verbal constraint, yet both are involved in the sentence material used. To distinguish between the effects of the two is of more than pedantic interest; it might increase our imperfect understanding of disturbances in verbal behaviour such as *receptive aphasia* where the patient's appreciation of pattern in what is spoken to him seems to have broken down. Unfortunately, it is impossible to separate syntax from verbal association in normal connected discourse, the two are so intimately

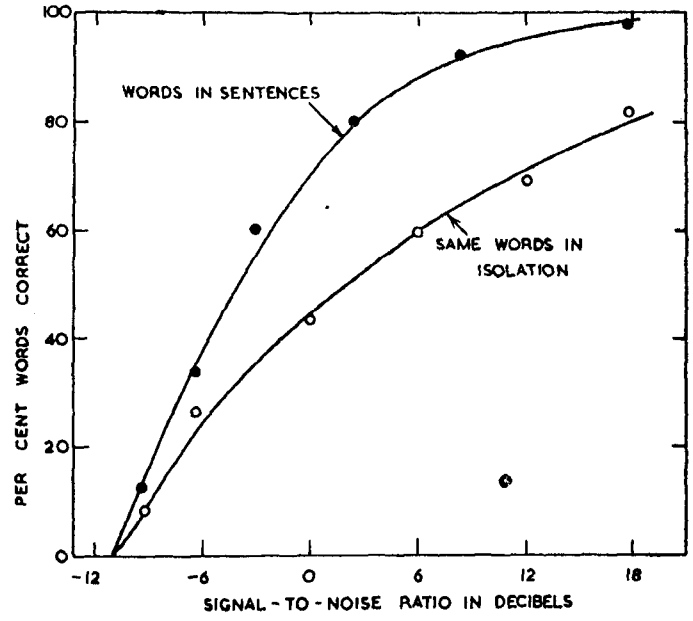

Fig. 1. Effect of sentence context on the intelligibility of words.*

fused. An attempt is being made at the University of Reading to study the problem through the use of an artificial 'language' which will allow such experimental interference. However, the results of another series of tests[3] do seem to show that verbal association alone is a strong contributory factor to the intelligibility of what is heard.

For this experiment, four sets of monosyllabic words were made up, matched as far as possible, in features such as frequency of occurrence, phonetic difficulty and so on. Subjects were given no information about the material except that it did consist of words, and they were asked to record what they

**From Miller, Heise, and Lichten, 1951. By courtesy of the Journal of Experimental Psychology and the American Psychological Association.*

thought they were. The words were heard, of course, in the presence of noise. As you may have anticipated, there was a difference between the sets. Two of them were random selections, while each of the other two consisted of words relating to a particular field of experience, actually 'food' and 'parts of the body'.

Subjects' scores on the two random series were so similar that we were able to average them, and in the graph below you can see a comparison between the average performance on the random series and that on words relating to parts of the body. The results are much like those obtained in Miller, Heise, and Lichten's experiment. Summing up in the same terms as before, the constraint of this type of verbal association enables the listener to hear about 25 per cent more of what is said to him at a signal-to-noise ratio comparable to conditions in a suburban street.

In this study we tried to get some information about how the

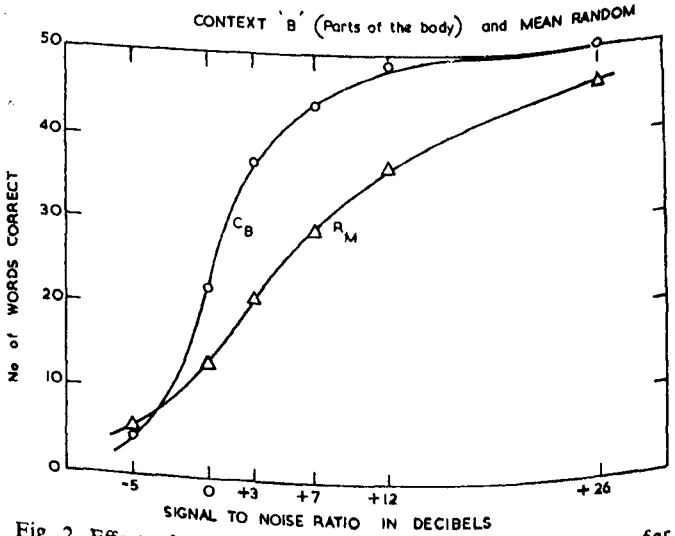

Fig. 2. Effect of verbal association upon intelligibility scores for words. (*Adapted from Bruce, 1956.*)

listener made use of contextual cues. Subjects were requested to describe their experiences after each trial, and their comments were often most revealing when related to their scores. For example, one subject commented at the end of the second trial that 'there seemed to be a relationship when several parts of the body were named'. She had rejected this theory (actually correct) because 'there were too many strangers' (her own incorrect responses). In doing so she unwittingly denied herself the use of the contextual aid present, and as a result achieved a score of 4 per cent correct compared to the group's average for the trial of 44 per cent correct (many of the subjects were already suspecting the constraint on the material at this stage and using their suspicions). However, at the end of the next trial she described her impressions in this way: 'So they are parts of the body! "Strangers" now rejected, whole words listened for instead of sounds.' The level of noise was reduced from trial to trial, so we should expect her score to rise, but the advantage gained from the use of context can be seen by once more comparing her score with the group average for the trial. The figures were respectively 58 per cent and 74 per cent correct. She had 'caught up' on the group average by 24 per cent of the experimental material!

The contribution of verbal context can be shown in another way. Fluent speaking and appreciation of speech demands that we should remember what has already been said. The constraints I have outlined would be of little use if the items conveying them, words, gestures and so on, were forgotten as soon as registered. We can see this most vividly in cases of *receptive aphasia*, where the patient flounders because he cannot apparently remember how he started a sentence. In the experiment to be described[4] a device is used that may serve as an introduction to the way in which questions of meaning are skirted, legitimately, in a field that would seem to necessitate their consideration.

As a starting point, we must accept the fact that meaningful verbal behaviour is familiar verbal behaviour. If I say to you 'blog', that remark has no meaning for you. But if I continue to say it, and you notice the specific occasions on which I use it,

then it comes to have meaning for you. I might be using it to represent a house, or a horse, or some thing or idea for which we haven't a word in our language. It doesn't matter – through repeated exposure to the sound and its situational context you understand me and the word is given meaning.

What is true of the single word is also true of sequences of words. The meaningful sentence is so because it is a familiar word pattern. We had a case of this before in 'Cat eats mouse' versus 'Cat mouse eats'. Now there is a simple way of measuring familiarity of verbal material. The more *frequently* a word or word sequence is used in our language, the more familiar it is to us. What happens if we make up word sequences whose character is increasingly constrained by frequency of usage? They become more meaningful. Consider these from Miller and Selfridge's experiment. First, *zero-order approximation.* 'Betwixt trumpeter pebbly complication vigorous tipple careen obscure attractive consequence expedition pane unpunished prominence chest sweetly basin awoke photographer ungrateful.' This is a statistical approximation to English made up by just drawing words at random from the dictionary. Any and every word had an equal chance to appear in the sample and there is no use made at all of the relative frequency of occurrence of the words in our language. It is called a zero-order approximation because no frequency restriction was placed on the choice of words. Then *first-order approximation.* 'Tea realizing most so the together home and for were wanted to concert I posted he her it the walked.' This isn't much better, but one gets the feeling that it is not quite so wild as the first example. It was constructed by picking words at random from continuous text, but giving each word a chance to occur proportional to its relative frequency in normal English – hence 'first-order approximation'. Going one stage further is *second-order approximation.* 'Sun was nice dormitory is I like chocolate cake but I think that book is he wants to school there.' In this example the relative frequency of word pairs is taken into account. Even in such a simple approximation as this you will notice that meaningful runs of words appear in the sample. It was obtained by

giving a person one word to use in a sentence, recording the word he used immediately after the test word, giving that as the test word to the next person, and so along the chain. The method is very convenient and can be extended to word triplets, quadruplets – up to whatever level of approximation one wishes, in fact. Thus: *seventh-order approximation.* 'Easy if you know how to crochet you can make a simple scarf if they knew the colour that it.' Using the relative frequency of seven-word units we obtain a sample which is almost completely meaningful. A tenth-order approximation would be even more like normal speech.

This relation between frequency of occurrence and meaning allows the worker interested in speech perception to pursue his experimental line without getting involved in semantic arguments. But to return to the relation between memory and verbal context. The increasing order of approximation entails an increasing degree of constraint, and this constraint is primarily of the sort I have described as verbal context. How well do people remember the different approximations? From the graph of Miller and Selfridge's results you will see that recall improves progressively with the order of approximation to normal language, i.e., with an increasing degree of contextual constraint. If we take a sample ten words long, the subject can remember almost 50 per cent more of a fourth-order approximation than a zero-order one. The graph also includes recall scores for text, that is, ordinary 'meaningful' written material. It is a striking feature of the results that on the whole the recall scores are as good for fourth- or fifth-order passages as they are for normal language. This is further support for the opinion that it is unnecessary to discuss problems of speech perception at a 'semantic level'. As Miller says, the results suggest that meaningful material is easy to remember, not because it expresses some undefinable concept or idea, but because it preserves the short-term dependencies of English word order. Verbal association and verbal context are then of great help in *remembering* what we are saying and what we are listening to.

Because of its lack of definition, I have left the discussion of

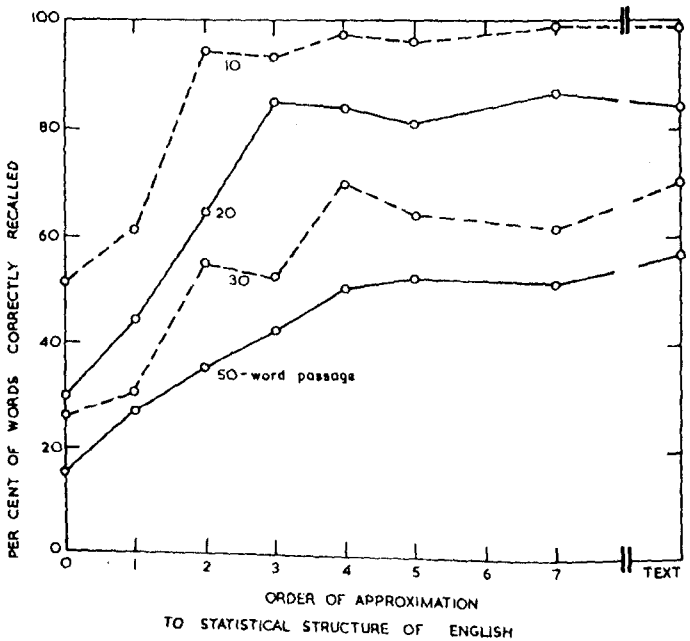

Fig. 3. Recall scores for sequences of words at different orders of approximation to English. Length of the sequence is the parameter.*

'situational' context to last. In listening to a spoken message we take advantage of a wide variety of cues over and beyond what is actually said. For convenience, we may include anything from facial expression to prior knowledge about the nature of the message. We can think of this type of constraint as a control which sets us to receive a particular class of message before the speaker opens his mouth, though it is also important for adjustments during the course of the message. If the speaker is white with rage, we already know the sorts of utterance he is likely to make. If we are at a football match, it is extremely unlikely

**From Miller and Selfridge, 1950. By courtesy of the American Journal of Psychology.*

that we shall hear a discussion on metaphysics. We gain a great deal of information about what is or is not likely to be said from such extra-verbal factors.

'Where so many individual features are concerned, it is difficult to be succinct. Perhaps some idea of the force of this constraint comes out in the last experiment to be described, though it must be noted that this test used a verbal short-cut to establish prior expectations in the listener.

Five standardized sentences were spoken to the subject under noisy conditions. They were repeated five times in different orders, but the subject did not realize that they were the same sentences. Before hearing each set of five, subjects were told that they would be given a key-word that might help them in hearing the sentences. In each case, the key-word was actually appropriate to only one of the sentences. The instructions and the key-words were given in the absence of noise, and the object of the test was to see what effect the attitude engendered by the key-word would have on the perception of the sentences. Here is a sample of one subject's attempted reproduction of the same sentence under the influence of three of the key-words. The key-words are given in order of presentation. *Test sentence:* 'We will give a prize to the top boy of each class.'

KEY-WORD	SUBJECT'S REPRODUCTION
Sport	'Football.'
Education	'He's top boy of his class.'
Weather	'We were ill advised to —.'

Although he has interpreted the sentence very nearly correctly under the appropriate key-word *education*, the next time the subject hears it he is off on a completely different tack. With the situation apparently one of *weather* reference, the same sounds lead to an emphasis on a different part of the test sentence, and a new interpretation, appropriate but incorrect. Sometimes the effect of this constraint took the form of a subject hearing practically the whole sentence correctly when it was preceded by the appropriate key-word, and then, on its repetition with an inappropriate key-word, hearing nothing intelligible at all. This

could happen even when the repetition followed the successful interpretation immediately!

The implication of all these experiments is the same: conditions for efficient speech transaction are those that foster active anticipations in speaker and listener. These anticipations are built upon familiarity with accepted verbal patterns, and intelligibility is in large part a function of this familiarity. This may sound a truism, but its recognition is important. It holds not only for the obvious cases of native versus acquired language, but for the ordinary everyday material of communication. There are today many situations in our mechanized world where such communication is necessary but difficult, and the resulting problems may best be solved by language design rather than by attempts at noise restraint. Also, communication has long outgrown simple face-to-face contact. We communicate over vast distances and by ever subtler means. The mechanical systems we employ are impressive but far from perfect. Immediate improvements are possible, not only through physical design but by language engineering. In its simplest form, language engineering is the design and selection of verbal material appropriate to a given task. In practice it ranges from choosing a standardized vocabulary for use in noisy plant to constructing an international language. The object is the same for the largest and the smallest case. It is to achieve intelligibility.

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FIRST TRANSATLANTIC TELEPHONE CABLE

ARCHIE CLOW

IN the last week of January 1957 nearly a thousand engineers, assembled in the Institution of Electrical Engineers in London, heard the meeting in which they were to participate opened by the President of the American Institute of Electrical Engineers speaking in New York. First, President Coover greeted the President of the Engineering Institution of Canada, who was presiding at a similar meeting in Montreal, and then Sir Gordon Radley, Director-General of the British Post Office, who was in the chair in London. For the first time, simultaneously in three world centres, a joint meeting was being held. Appropriately it was a symposium on the design, manufacture, and installation of the first transatlantic telephone cable, because it was by using the new cable that such a meeting became a possibility. True, a radio-telephone link could have been used in the past, but the magnetic disturbances to which the North Atlantic is prone might have made impossible any meeting planned in advance. With the new cable, however, communication can be maintained with a quality comparable with an inland trunk call.

Electrical communication across the Atlantic Ocean is just over ninety years old. In 1866, after a number of abortive and costly failures, a successful telegraph cable was completed between the south of Ireland and Newfoundland. Since then, submarine telegraph cables have been laid in all the oceans of the world. There is, however, one signal difference between these telegraph cables and the first telephone cable. This is the inclusion, as an integral part of the scheme, of submerged flexible amplifiers (repeaters) throughout the long and deep transoceanic section.

Objectives and Route

The project was planned to link the vast telephone networks of Britain and continental Europe with those of Canada and the United States through London and New York, with a branch to Montreal. From the point of view of new achievement in electrical engineering the main points of interest are the two submarine sections, the first across the Atlantic from Scotland to Newfoundland, and the second from Newfoundland to Nova Scotia. Each of these called for the solution of problems determined by the length of the link, the depth of the ocean that it had to traverse, and the bandwidth demanded by the telecommunication engineers.

The circuits between London and New York can be divided into five sections. (*a*) The first part of the route is from London to Oban, a distance of 561 miles. As far as Glasgow the circuits are carried on a 24-circuit carrier cable; from Glasgow to Oban a new coaxial cable, some 111 miles in length, was specially laid. (*b*) Between the Oban terminal station in Scotland and the Clarenville station in Newfoundland, a distance of 2,240 miles, two submarine cables were laid. The reason for the two cables is that the flexible amplifiers incorporated in the deep-ocean cable (some of it lies over two miles deep) can only carry speech in one direction. Fifty-one of these amplifiers are incorporated in each cable, thus permitting the transmission of three 12-channel groups. (*c*) The next part of the route (Clarenville to Sydney Mines: 376 miles) crosses both wild Newfoundland terrain and the comparatively shallow continental shelf round the Canadian coast. This permits of the use of 'rigid' amplifiers of British Post Office design. These can transmit speech in both directions so a single cable is adequate. (*d*) Between Sydney Mines and Portland, Maine (575 miles), connexion is made by a radio-relay system with 17 intervening relay stations. Three hundred and twenty-eight miles along the route there is a break-out point for the Canadian circuits. (*e*) Beyond Portland, use is made of the American Telegraph and Telephone Long Lines network. In all, some 4,500 miles of submarine cable with 118 submerged amplifiers are involved. Cables and repeaters provide a usable bandwidth of 144 kc/s across the Atlantic, and 240 kc/s between Newfoundland and the Canadian mainland. The transoceanic link thus determines the overall circuit provision. These are divided as 29 circuits between London and New York, and 6 between London and Montreal. Of the New York circuits 7 are extended to continental centres – to Paris, Frankfurt, Amsterdam, Brussels, Copenhagen, and Berne. The longest link, New York to Copenhagen, is nearly 5,000 miles long. The 36th circuit is divided for narrow-band use between London and New York, and London and Montreal.

British and American Prototypes

The position of the British Isles on the continental shelf determined the nature of the work undertaken by the Port Office in establishing contact with the Continent by submarine telephony. This was achieved by short, shallow-water cables, at first without amplification. It was obvious, however, that the incorporation of amplifiers would be an advantage, and the first 'repeatered installation' was laid between Wales and the Isle of Man in 1943. The amplifiers used were large rigid structures, but they had the advantage that they could handle two-way traffic, 'go' and 'return' messages being carried by different frequency bands.

With the experience of the Isle of Man cable behind them, the Post Office began a design study for a deep-sea cable in 1948. This led in 1952 to a project for a cable between Aberdeen, Scotland, and Bergen, Norway – a transoceanic prototype, and the longest submarine cable then projected.

American telecommunication engineers had been considering the possibility of a telephonic link with Europe for over thirty years, but nothing had come of earlier suggestions on account of the economic depression of the early thirties and the advent of radio-telephony, which initially met current demand for transoceanic communication. Technological exploration continued, however, and the Bell Telephone Laboratories conceived the idea of using two cables incorporating, virtually as part of

the cables, flexible amplifiers of simple design and maximum reliability – with an expectation of life of twenty years or more. As a prototype for a transatlantic cable, a system working on the above principles was opened between Key-West and Havana. This involved the use of two cables, some 120 miles long, each with three flexible repeaters working at a depth of nearly a thousand fathoms.

This was the background of experience on the two sides of the Atlantic when representatives of the British Post Office and the American Telegraph and Telephone Company met in 1952. Each had its own peculiar experience to contribute, Britain to shallow-water and America to deep-water problems. Based on this, Dr Mervyn Kelly, for the Bell System, and Sir Gordon Radley, for the Post Office, recommended that the American system be used for the long deep Atlantic crossing, and the British Post Office system for the Newfoundland/Nova Scotia link.

The Atlantic Crossing

It is just ninety years since the first successful transatlantic telegraph cable was completed by the *Great Eastern*. This ran from Valencia, in Ireland, to Trinity Bay, Newfoundland. Since then, many more cables have been added: fifteen direct, and five by way of the Azores. Thus the problem of selecting the route for the first telephone cable was no simple one. Various factors that might lead to an interruption of the service had to be taken into account. A faulty cable has to be brought to the surface by grappling, thus it is highly desirable to avoid existing cables – for mutual safety if nothing else. In the past, trawling has been one of the main causes of interruption in the submarine cable service, so the avoidance of fishing grounds is also a major consideration. The same is true of anchorages frequented by merchant shipping. As far as sites for terminal stations are concerned, problems of accessibility, suitability of shore approach for the cable, amenities for staff, and avoidance of potential military target areas, had all to be considered.

With regard to the cable itself, the shorter the route the better.

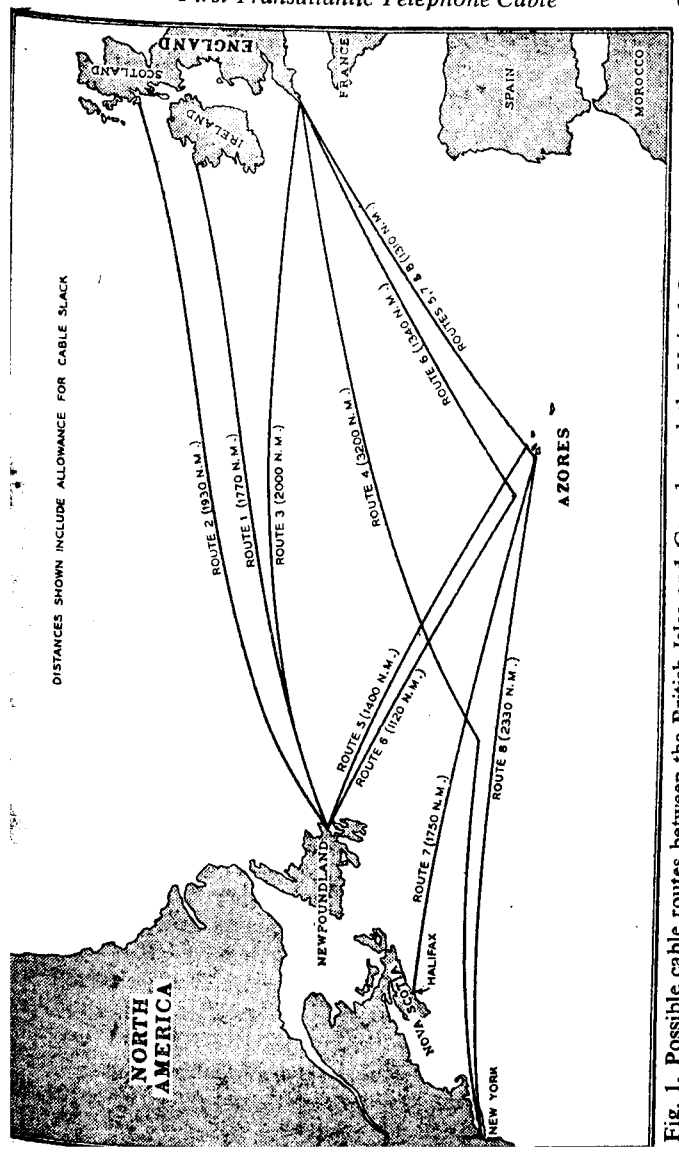


Fig. 1. Possible cable routes between the British Isles and Canada and the United States. (*Courtesy of the Institution of Electrical Engineers.*)

This eliminated the idea of a direct route from Cornwall to New York. Routes via the Azores had certain attractive features counterbalanced by the undesirability of the cable crossing foreign territory *en route*. Cornwall to Newfoundland was of approximately the right length, but the congestion of existing cables at the eastern end was considered to present too great a hazard to so important a link.

The final choice was Newfoundland to Scotland, subject to a suitable terminal site being found there. It was known that the Firth of Lorne, between the Island of Mull and the mainland, was a quiet channel, and several small bays south of Oban were examined by a survey party. One, three miles south of Oban, called Port Lathaich, was selected. Figure 1 illustrates some of the possible routes that were considered.

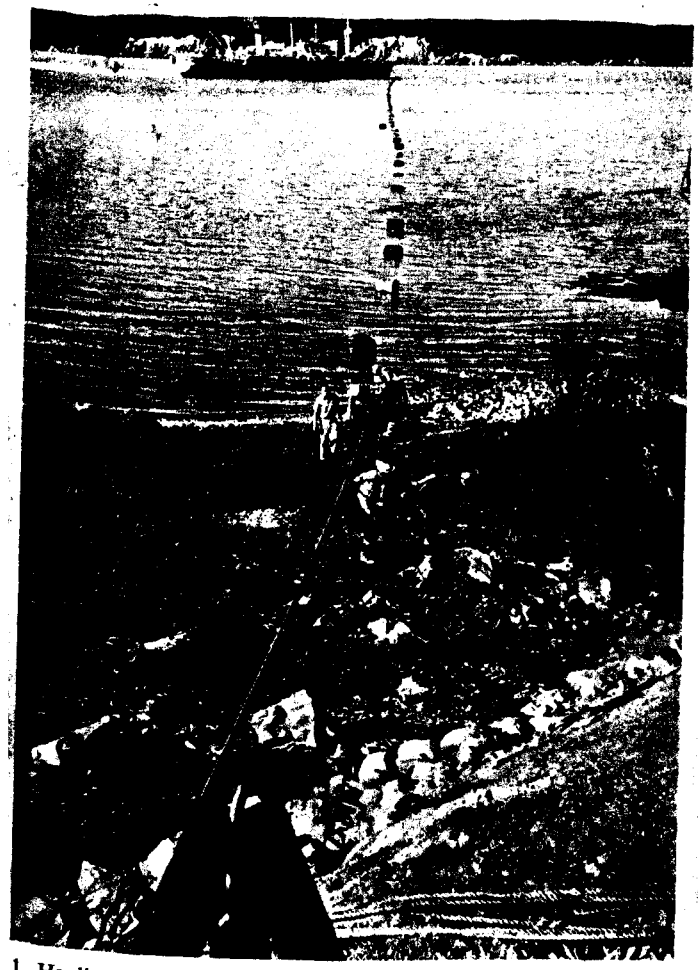
On the Canadian side there was no workable alternative to taking the cable into Trinity Bay, the same bay into which the first telegraph cable was brought in 1866. A careful survey led to the selection of Clarenville, a small town on the railway, as the best site for the landing and the terminal station (Inset 1).

The route between Oban and Clarenville enabled the new cable to be laid clear of all existing cables, and avoiding trawling areas and anchorages (Insets 2, 3, 4, and 5). The two terminal stations are within easy reach of towns and meet the requirements of accessibility and strategy. Figure 2 illustrates the transatlantic profile of the route finally selected.

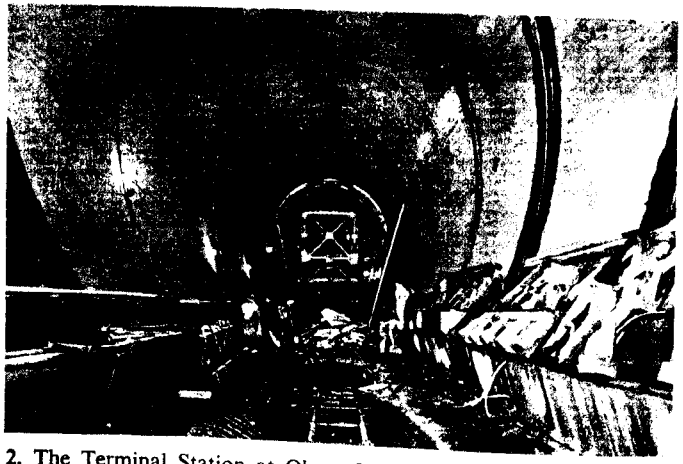
Across Newfoundland and the Cabot Strait

While the North Atlantic crossing is undoubtedly the most dramatic part of the enterprise, considerable thought had to be put into working out a route between Clarenville and a point accessible to the North American continental network. Several schemes were considered: that chosen involved an overland cable from Clarenceville to Terrenceville, on Fortune Bay, and then a direct submarine cable to Sydney Mines, Nova Scotia. Part of the 'overland' route is through lakes, thus saving the cost of trenching, and the submarine part across Cabot Strait is clear of other cables. Standard solid-dielectric coaxial cable is used

THE FIRST TRANSATLANTIC TELEPHONE CABLE

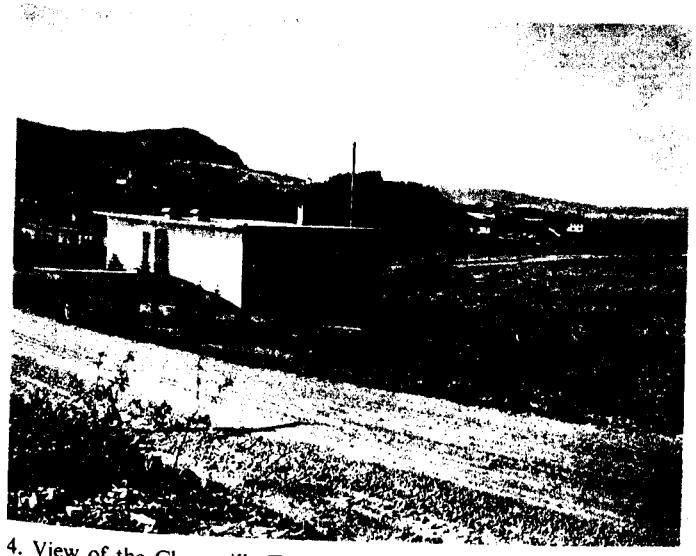


1. Hauling the Canadian end of the Transatlantic Telephone Cable ashore from H.M.T.S. *Monarch* at Clarenville, Newfoundland. (*Courtesy of H.M. Postmaster-General*)

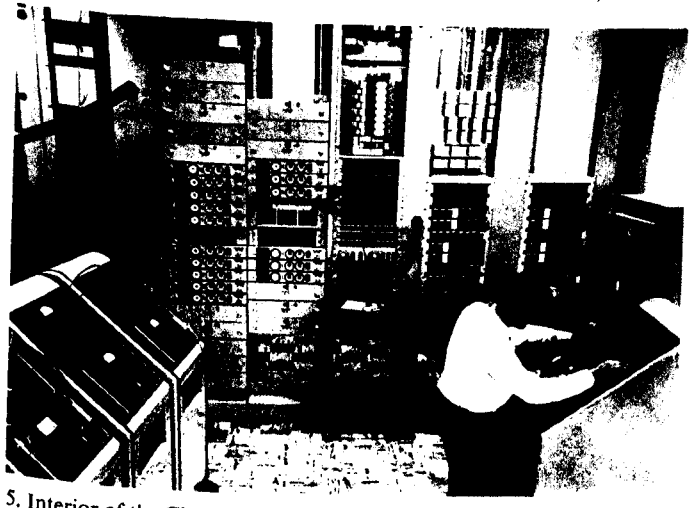


2. The Terminal Station at Oban, Scotland: view of the apparatus tunnel which is hollowed out of a rocky hillside. (*Courtesy of H.M. Postmaster-General*)

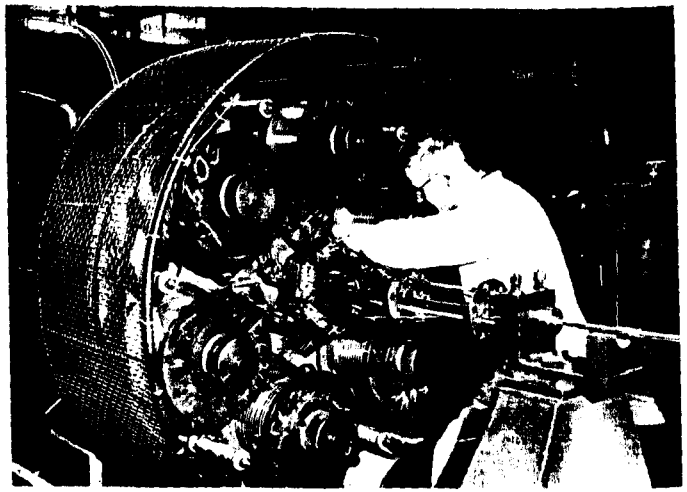
3. Oban Terminal Station showing an engineer testing at the frequency base unit in the telephone repeater room. (*Courtesy of H.M. Postmaster-General*)



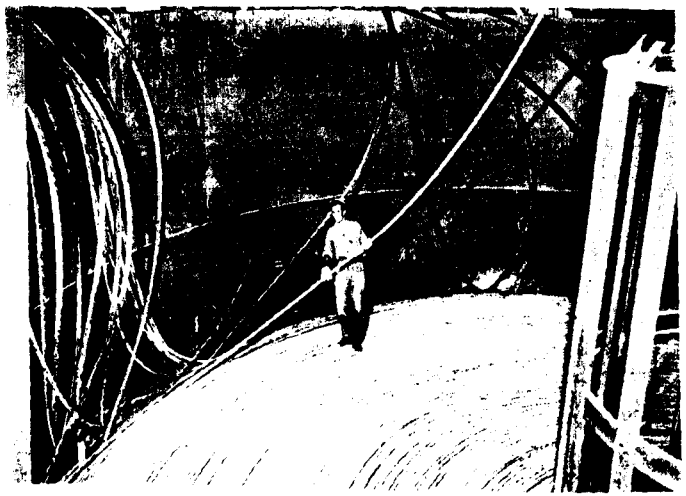
4. View of the Clarenville Terminal Station on the shores of Trinity Bay, Newfoundland. (*Courtesy of H.M. Postmaster-General*)



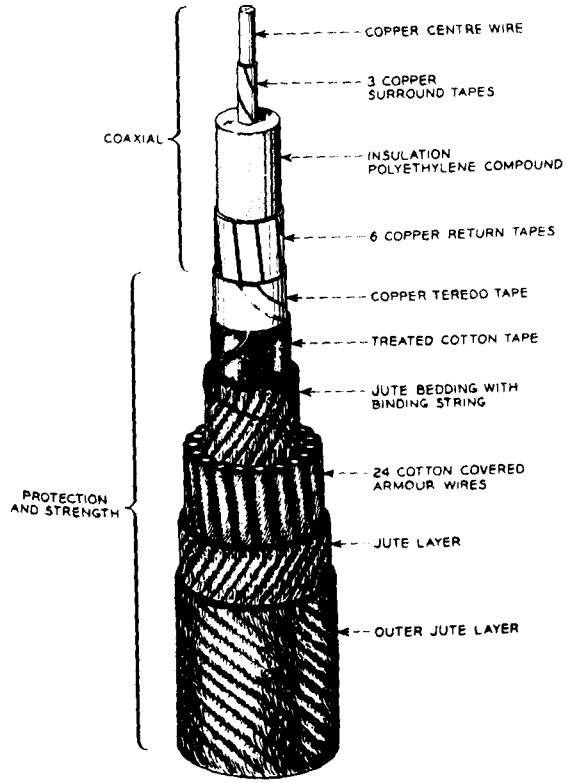
5. Interior of the Clarenville Terminal Station with an engineer engaged in checking equipment. (*Courtesy of H.M. Postmaster-General*)



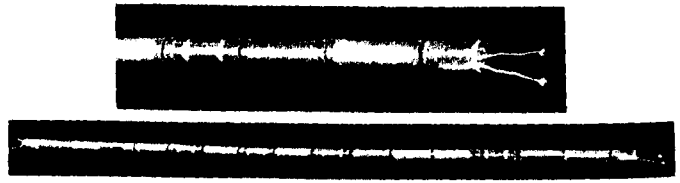
6. Manufacturing the cable: copper serving and taping machine at the works of Submarine Cables Ltd, Erith, Kent, England. (*Courtesy of H.M. Postmaster-General*)



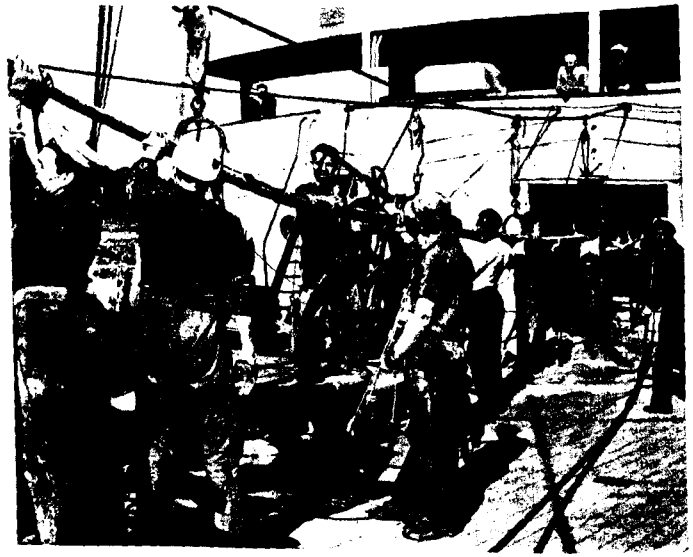
7. American manufactured cable stored in giant tanks at the Simplex Wire and Cable Co's works at Newington, New Hampshire. (*Courtesy of H.M. Postmaster-General*)



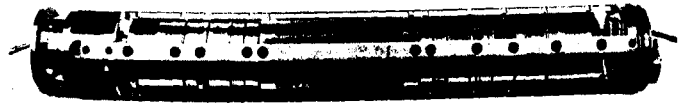
8. Detail of the construction of the Transatlantic Telephone Cable. The whole of the cable is of substantially the same design throughout except that the degree of protection varies with the position of the cable: it is greater at the shore ends than in mid-Atlantic. (*Courtesy of the Institution of Electrical Engineers*)



9. Overall view of the flexible repeaters designed for the main deep section of the Atlantic crossing. (*Courtesy of the Institution of Electrical Engineers*)

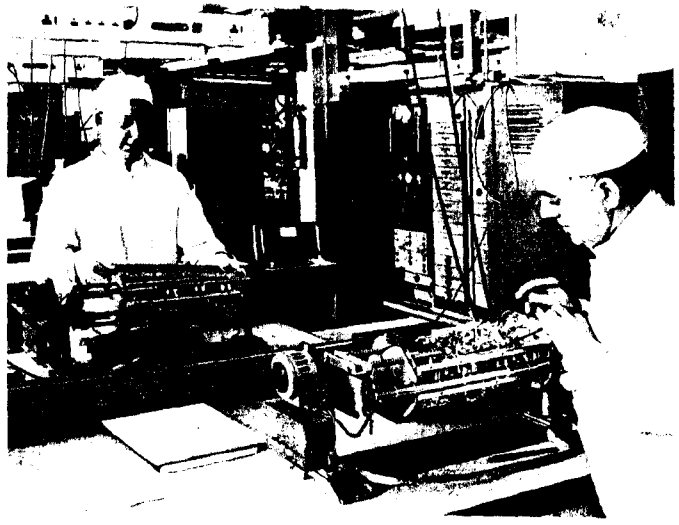


10. A flexible repeater, jointed into the cable, being loaded into the storage tanks of H.M.T.S. *Monarch*. (*Courtesy of H.M. Postmaster-General*)

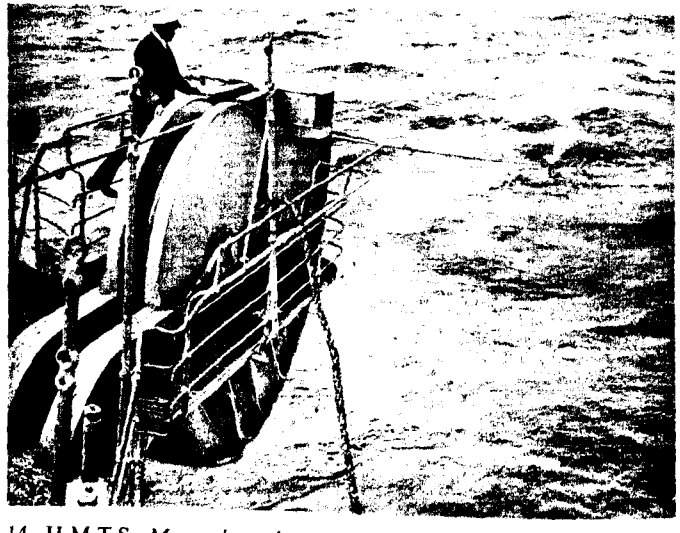


11. British Post Office rigid repeater with the external watertight casing removed. Reading from left to right the various components include a resistance unit, power separating filters, bridge and equalizer, directional filter, amplifier, second directional filter, second bridge and equalizer, supervisory unit, further power separating filters, and lastly a moisture detector. (*Courtesy of H.M. Postmaster-General*)

12. Wiring up directional filters for the rigid repeaters in the Clarenville-Sydney Mines section of the system. (*Courtesy of H.M. Postmaster-General*)

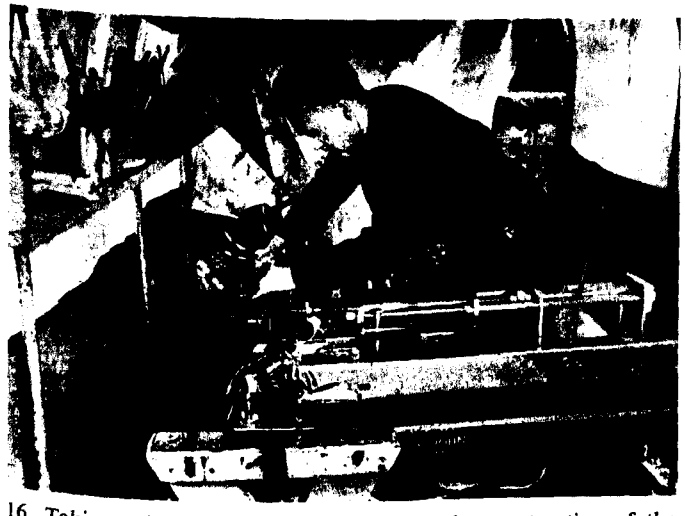


13. Completing a rigid repeater amplifier unit. (*Courtesy of H.M. Postmaster-General*)



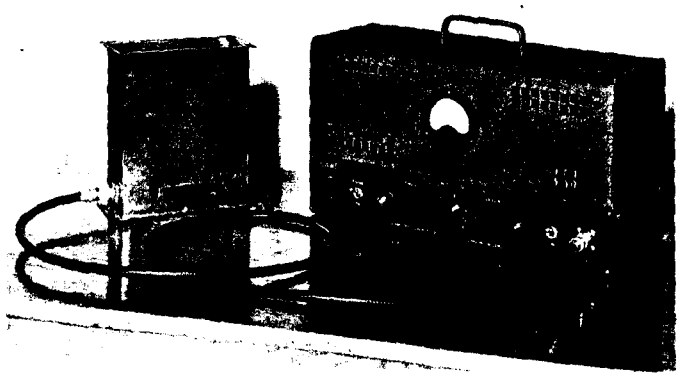
14. H.M.T.S. *Monarch* paying out cable astern. (*Courtesy of H.M. Postmaster-General*)

15. An example of the rugged Newfoundland terrain across which the cable had to be laid. (*Courtesy of H.M. Postmaster-General*)

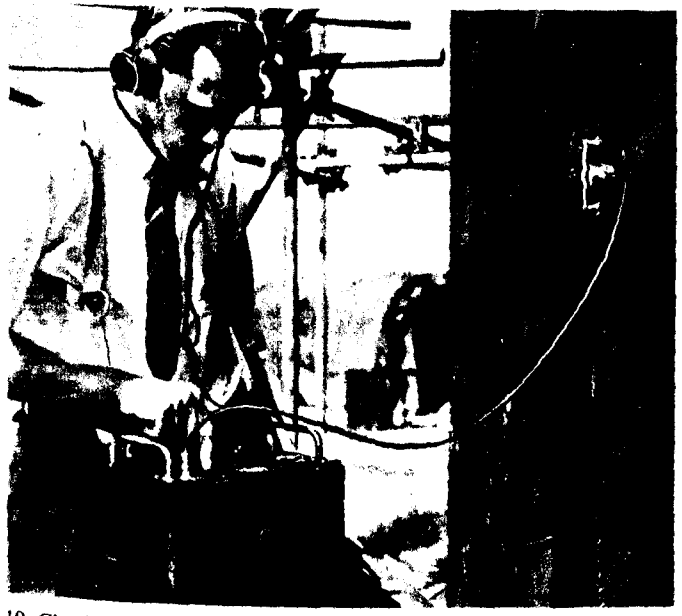


16. Taking a joint out of a mould during the construction of the Newfoundland link. (*Courtesy of H.M. Postmaster-General*)

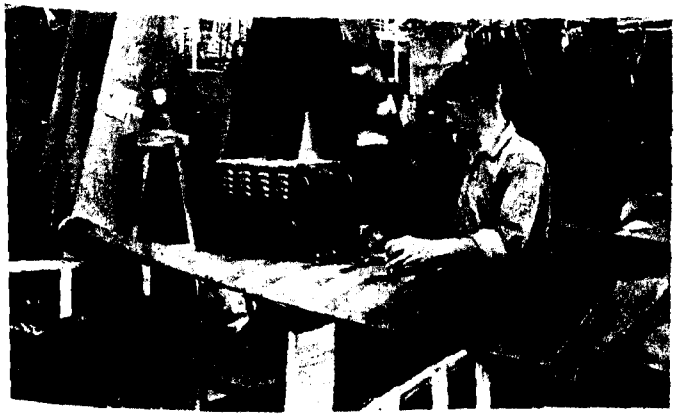
17. Treating the surface of a large aluminium casting with an ultrasonic soldering iron. (*Courtesy of Mullard Ltd*)



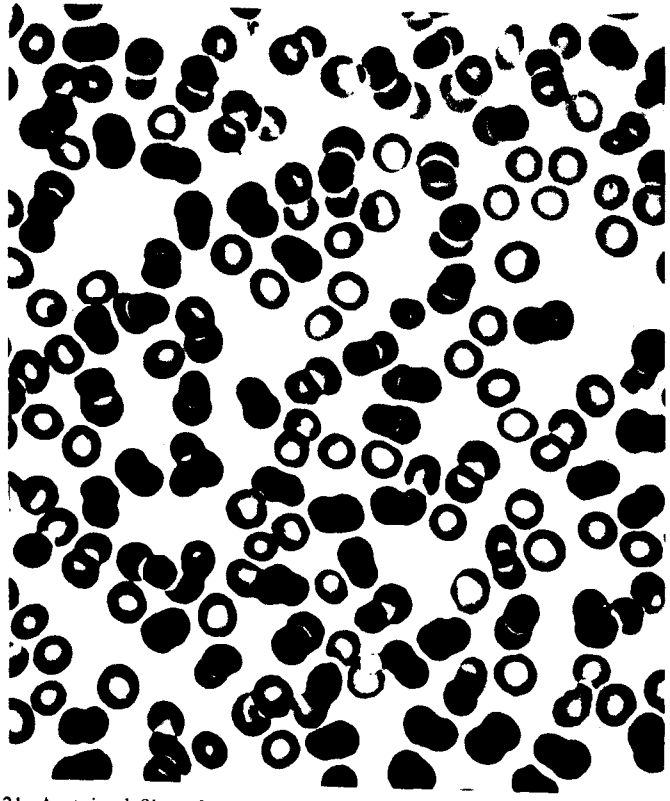
18. Cleaning by Ultrasonics: cleaning tank with built-in transducer and control gear. (*Courtesy of Dawe Instruments Ltd*)



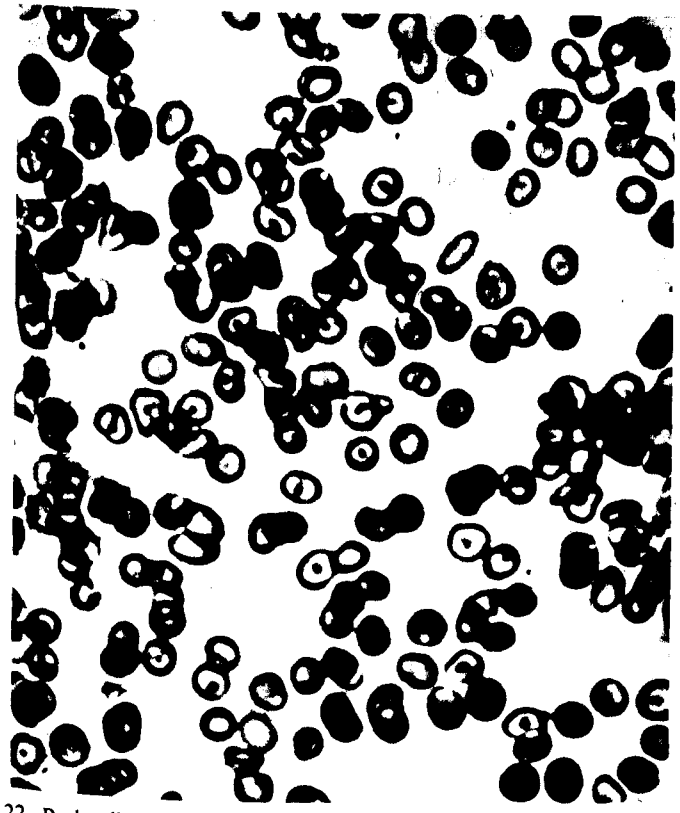
19. Checking the thickness of a pressure vessel in service with a portable ultrasonic gauge. The transducer is being held in position by magnets. (*Courtesy of Dawe Instruments Ltd*)



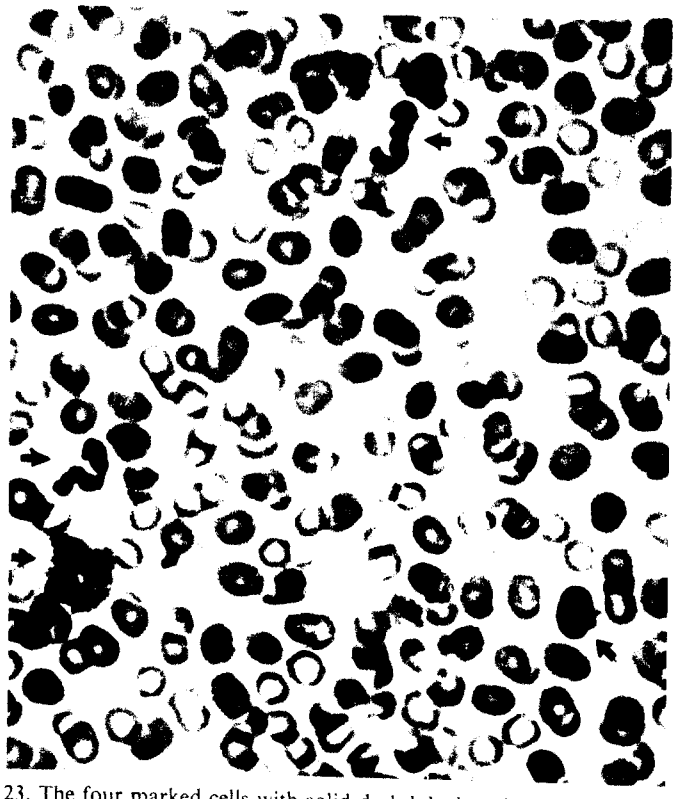
0. Checking propeller blades made from hollow steel tubes with an utomatic ultrasonic gauge. (*Courtesy of Dawe Instruments Ltd*)



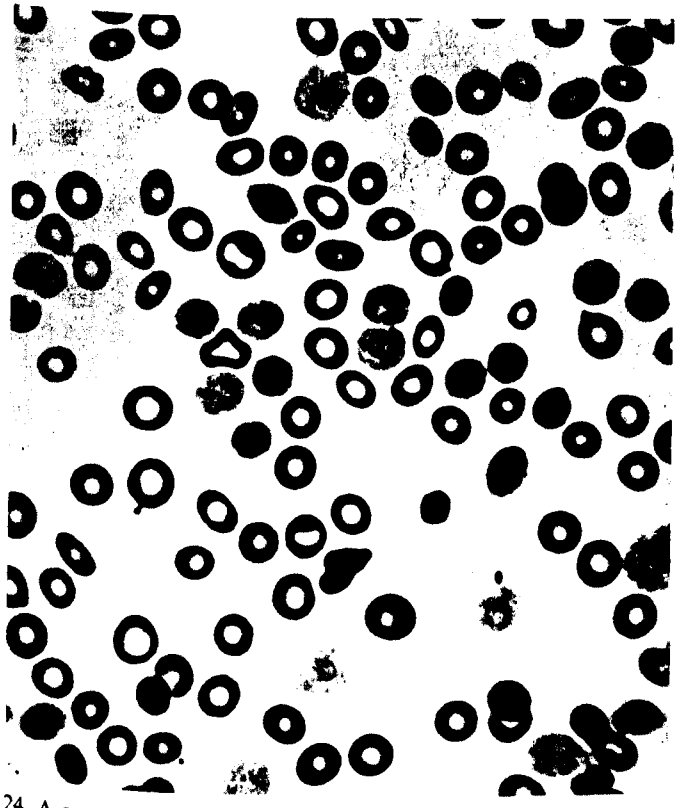
21. A stained film of normal red cells. The erythrocytes are all seen from above since they are flattened on to the glass slide. (×720)



22. Red cells in a case of iron deficiency anaemia. Apart from the abnormal shapes there are many cells that show a central deposit of haemoglobin: these are called 'target' cells. Their presence is sometimes associated with an inborn error of haemoglobin metabolism. (×720)

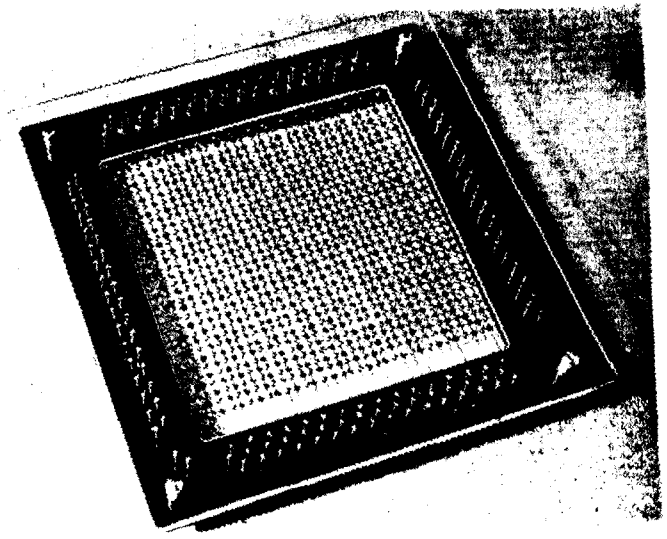


23. The four marked cells with solid dark lobed nuclei are leucocytes. From a normal subject. (×720)



24. A comparable field from an individual with lymphatic leukaemia. The rounded solid mottled cells are leukaemic lymphocytes, while the rather larger smudges are effete cells damaged in the preparation of the film. (×720)

25. A ferrite rod aerial of the type suitable for radio receivers. (*Courtesy of Mullard Ltd*)



26. A magnetic 'memory' matrix using small rings of a 'square loop' ferrite material. This is a new device for electronic digital computers capable of storing over a thousand bits of information. (*Courtesy of Mullard Ltd*)

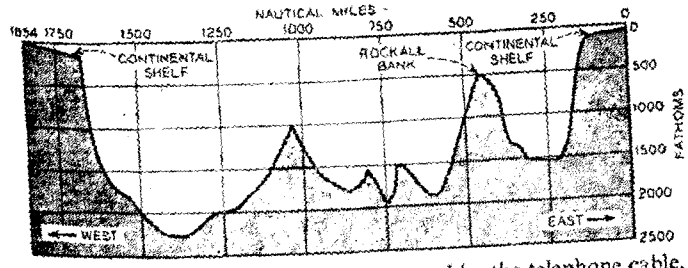


Fig. 2. Profile of the Atlantic Ocean crossed by the telephone cable.

throughout, and, both on land and in the shallow waters of the Strait, amplifiers of Post Office design are used. Such a shallow-water cable is, however, more prone to interference, so there is the difference that iron shielding tapes and a plastic jacket had to be added to the standard cable. Beyond Sydney Mines the route is carried forward by radio links that do not concern us here. (Insets 15 and 16.)

Submerged Repeaters

As has been indicated earlier, the feasibility of the whole enterprise depended on the possibility of designing, constructing, and laying amplifiers that might be expected to have an undisturbed life of not less than twenty years.

Experiments with this end in view were begun in America at the end of World War I, under the aegis of Dr O. E. Buckley, O. B. Jacobs, and J. J. Gilbert. Their objective was to design a thin flexible unit, which, spliced into the cable, would be capable of passing from the cable stores on board ship directly through the conventional paying-out gear and so uninterrupted to the ocean bed. (Insets 10 and 14.) Practical achievement came in 1950 with the laying of 120 miles of repeatered cables, at a maximum depth of 900 fathoms, between Key-West, Florida, and Havana, Cuba. (Inset 9.) Experience gained in this operation formed the basis of the transatlantic design. As developed, the American repeater appears as a flexible bulge in the cable, some eight feet long and less than three inches in diameter. At

either end the bulge tapers down to cable diameter over a distance of twenty feet.

Because of the very restricted space for components inside these repeaters, only simple circuits can be used. The repeater consists, therefore, of a conventional 3-stage feedback amplifier without duplicated components. Moreover space is not available for directional filters so the amplifier works in one direction only, and for that reason two cables are required. As the repeaters must have a high degree of flexibility they are built up of 17 sections, linked mechanically with springs, and electrically by bus tapes. Each section consists of a circuit component, or group of components, mounted in methyl methacrylate plastic, selected for its desirable physical and chemical properties.

In order to achieve the extreme reliability and stability that are essential, and to make as certain as is humanly possible that there would be no failures of the repeaters during their estimated period of service, they were manufactured in a special building at the Western Electric plant, at Hillside, New Jersey. The rooms in which they were assembled were air-conditioned, kept at constant temperature and humidity, and at a positive pressure so as to avoid the ingress of dust into the 'restricted' area. Nylon clothing and special shoes were issued to personnel. The progress of the construction of each repeater was recorded on special dust-free paper. Each of the 6,000 components required had to pass the most stringent tests, as also had the completed repeaters. The glass and rubber seal assemblies, which have to withstand the great pressures of the ocean depths, had to pass a helium leakage test which would detect the passage of one cubic centimetre of helium in thirty years. As if this were not enough, it was supplemented with another radioisotope leak test.

In terms of the agreement the repeaters for the Newfoundland/Nova Scotia link were of Post Office design. They were made at the Standard Telegraph and Cables factory at Woolwich, London. There, as at Hillside, a special building, air-conditioned, and supplied with filtered air, was equipped for the purpose.

Designed for shallower water than the flexible repeaters, the Post Office repeater is contained in a rigid steel housing, some nine feet long and about one foot in diameter, and the additional space is utilized to accommodate, in addition to the components found in the American repeaters, the directional filters necessary for two-way amplification, and dual amplifiers, so arranged that any one component in one can fail without affecting the operation of the other. (Insets 11, 12, and 13.)

The Cable

The cable which is the lifeline of the whole project does not differ substantially from those already in use in existing shallow-water links except in two respects: the vital importance of the facility it provides, and its much greater length compared with any other existing telephone cable. These two aspects were all-important, however, and they led to the drawing up of the most exacting specification for a submarine cable that has ever been prepared. Over 4,000 miles of the cable were manufactured in a new factory built by Submarine Cables at Erith, Kent. (Inset 6.) The rest, less than 500 miles, was made in Essex, England, and New Hampshire, U.S.A. (Inset 7.)

The cable is coaxial. The composite central conductor is covered with polyethylene insulation, to form the core, on which is laid the return conductor of six copper tapes and a copper binder tape. This is lapped with impregnated cotton tape followed by a jute serving to form a bed for the steel armour wires, which are, in turn, covered with further jute servings and compounds. (Inset 8.)

The overall size of the composite conductor had to be controlled within narrow limits. Close to where the insulated conductor emerged from the extruder there was an electronic device which checked the diameter; another measured and recorded its electrical capacitance, and both these devices exercised servo control on the extrusion. A third device recorded concentricity of the centre conductor within the insulation. Many tests were applied as the cable was manufactured and stored: the application of 90,000 volts D.C. to detect any weak spots that might

develop into points of failure; the X-raying of every joint to see that it was perfectly executed and did not contain voids or foreign matter, and so on.

Conclusion

Speaking at the end of the historic meeting that linked the engineering institutions of Britain, America, and Canada, Sir Gordon Radley said: 'Immediately after the opening of the new service, the traffic to both New York and Montreal increased substantially and now stands at about twice the pre-cable level to New York, and about three times to Montreal. In the short time that has elapsed since the opening date this is how the public has reacted to a reliable high-grade telephone service across the Atlantic. How much further the traffic will increase remains to be seen, but it is already clear that we must consider the pattern of the next transatlantic telephone cable.'

NEW MAGNETIC MATERIALS THE FERRITES

F. BRAILSFORD

ALL material substances are magnetic in some degree, but it is mainly the strong and obvious phenomenon known as ferromagnetism that interests the technologist. Among the elements this useful property occurs only in iron, cobalt, nickel, and the rare element gadolinium. But ferromagnetism also occurs in many alloys, a few of which are composed wholly of non-ferromagnetic elements. Similar strong magnetic properties are found in the mineral magnetite, the lodestone of antiquity. Recently a new series of useful materials closely related chemically, but much superior to magnetite, have been produced artificially. These materials are known as ferrites.

In most electrical engineering equipment the magnetic materials employed have magnetism induced in them, and this magnetism undergoes continuous and rapid reversals. The number of complete cycles of this alternation of the magnetization occurring every second depends on the purpose for which the equipment is designed. In the magnetic core of a transformer connected to the British Grid, for example, the alternating frequency is fifty times every second. In a small audio-frequency transformer core frequencies of some thousands of cycles per second would occur. However, very much higher frequencies, rising indeed to values as high as 10,000 megacycles per second, are employed in the propagation of electromagnetic radiations in radio, television, and in radar. At these very high frequencies alternating magnetization cannot be usefully produced in the ferromagnetic metals because, since they are good conductors of electricity, internal parasitic electric currents of such magnitude are set up that overheating and other undesirable effects

occur. The ferrites, on the other hand, have the supreme advantage over metals that they are exceedingly poor conductors. They are less conductive than the metals by a factor of the order of 10^8. As a result, their magnetic properties can be used at very high frequencies without the overwhelming effects of internal 'eddy-currents'.

It is of some interest to recall that, in the year 1820, in a celebrated experiment, Oersted showed that there was a connexion between electricity and magnetism when he found that a pivoted compass needle was deflected by an electric current in a neighbouring conductor. A few years later Ampère proved that a circulating current in a closed circuit produced external magnetic effects identical with those of an equivalent magnet. Ampère went on to suggest that there was no magnetism except that which was due to electric currents. His view was that the molecules comprising a magnetic material were themselves elementary magnets due to electric currents permanently circulating in minute internal circuits. This is indeed also the modern idea, but it is now expressed in precise and detailed terms.

From the picture of atomic structures that has been built up by physicists we may, for present purposes, depict a single free atom of iron by the diagram in Figure 1. This shows a central nucleus surrounded by 26 orbital electrons arranged in sub-shells in the structure in accordance with the numbers in the diagram. Furthermore, each electron, negatively charged, spins like a gyroscope on its own axis and, recalling Ampère's hypothesis, thereby constitutes an electric current in a minute circuit. Because of the spin every electron in the atom has, in fact, a magnetic moment of unit amount known as a Bohr magneton. In each completely filled sub-shell, however, every electron is paired with another spinning on a parallel axis, but with spin in the opposite direction. These filled sub-shells therefore have no resultant magnetic moment. The incomplete 3*d* sub-shell, however, has five positive spins on parallel axes to only one negative, and the magnetic moment of the atom as a whole results from the four uncompensated spins and so amounts to four units. This is shown diagrammatically in the figure.

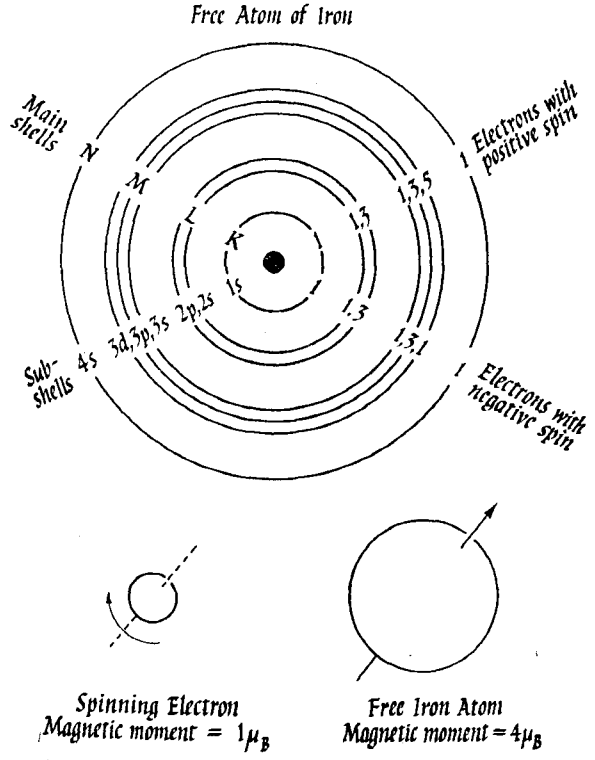


Fig. 1. Electronic Structure of a Free Atom of Iron.

At normal temperatures iron crystallizes on a body-centred cubic lattice as shown in Figure 2 (*a*). Now in any ferromagnetic crystal there are very powerful forces of interaction between the spinning electrons of the neighbouring atoms. In ferromagnetism this interaction is said to be positive, meaning that the forces hold the magnetic axes of the atoms parallel, with the magnetic moments pointing in the same direction. There are also other very much weaker forces present which are associated with the geometry of the crystal. In iron, in the absence of other external

forces, the result is that the atomic magnets all point together in one direction parallel to any one of the cube edges of the crystal. The possible directions for this spontaneous magnetic saturation are shown in Figure 2 (*a*). It is now well established that each crystal of the polycrystalline metal is also subdivided by boundary surfaces into small regions called domains. There are, as stated, six possible cube-edge directions in a crystal of iron for the domain magnetization, and although every part of the crystal is magnetically saturated, the crystal as a whole may show no external magnetic effect because the directions of the magnetic axes in the constituent domains are distributed in a random fashion amongst the six directions. In the core of a current-carrying coil a magnetic force is exerted upon the domains as in Oersted's experiment. This will produce a resultant magnetization in the direction of the force due in part to a rotation of the magnetic axes of the domains and also to movements of the boundary surfaces causing some domains to grow in size at the expense of others less favourably directed.

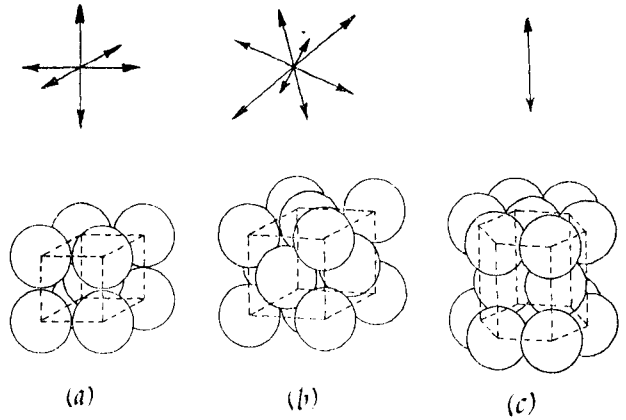


Fig. 2. Crystal Structure of (*a*) Iron – body-centred cubic, (*b*) Nickel – face-centred cubic, and (*c*) Cobalt – hexagonal. Possible directions for the spontaneous magnetization in the domains are shown above, namely, six cube-edged directions for iron, eight long-diagonal directions for nickel, and two axial directions for cobalt.

Nickel, on the other hand, has a face-centred cubic structure as shown in Figure 2 (*b*). The crystalline forces cause the domain magnetization to lie in the direction of a long diagonal of the cube. There are therefore in this example eight such directions as shown by the arrows. In cobalt, which is hexagonal (*c*), the equilibrium positions are axial, and so only two in number.

THE FERRITES

In its natural impure state the iron oxide Fe_3O_4, or magnetite, has some of the 'permanent magnet' attributes of steel in so far that it will attract pieces of iron and the like. In purified form it is more akin to soft iron which does not readily retain its magnetism. J. L. Snoek[1] carried out a systematic investigation of oxides of the same crystallographic structure as magnetite. He

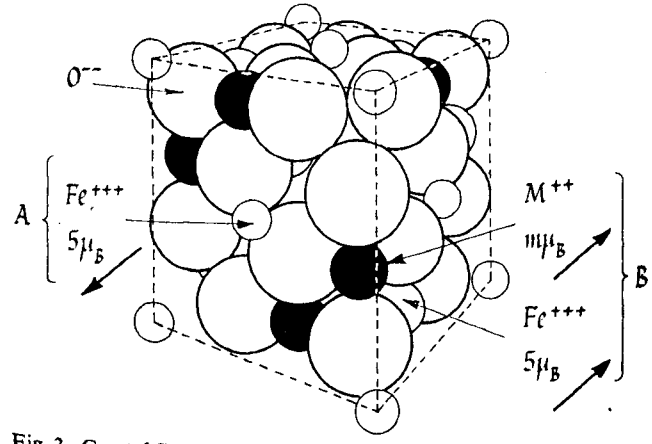


Fig. 3. Crystal Structure of Magnetic Ferrites (Magnetite, Fe_3O_4).

substituted other elements for some of the iron atoms in the material and thus created a series of hard, ceramic-like, and much-improved materials which are now known as ferrites.

Magnetite itself may in this connexion be called an iron ferrite. It is of ionic cubic crystal structure, as represented in Figure 3,

which shows a volume appropriate to eight complete molecules. The ions present in each molecule may be denoted by writing

$$Fe_3O_4 = Fe^{+++}\ (Fe^{++}\ Fe^{+++})\ O_4^{--}$$

Each oxygen ion receives two electrons from the iron present to form a negative ion O^{--} with two negative charges. One iron atom per molecule gives up two electrons to form the positive iron ion Fe^{++}, while the other two each contribute three electrons thus forming the trivalent ions Fe^{+++}. The large spheres in the diagram represent the oxygen anions packed together on a face-centred cubic crystal lattice. In the completely ordered crystal the smaller iron cations occupy regular cubic positions in the interstitial spaces between the oxygen ions. There are two types of available space for them, known as *A* and *B* sites. These sites differ in size and in the relation of the site to the surrounding oxygen ions. These differences are illustrated in Figure 4, where it is seen that the *A* site is the smaller. The iron ions on

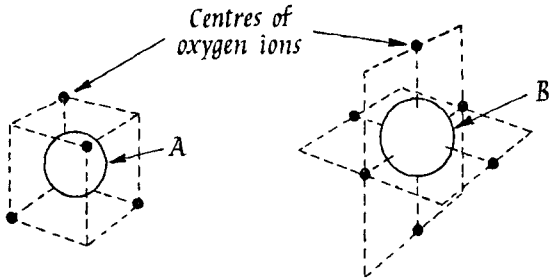


Fig. 4. Positive ions on *A* and *B* sites in relation to negative oxygen ions. *A*-ions have four near-neighbour oxygen ions in tetrahedral positions: *B*-ions have six near-neighbour oxygen ions in octrahedral positions.

A sites also have four near-neighbour oxygen ions while those on *B* sites have six. These site differences are important in the theory that explains the magnitudes of the spontaneous magnetic saturation values observed in different ferrites. In Figure 5 all the iron ions appropriate to eight complete molecules have been abstracted. On the left are trivalent ions on *A* sites. On the

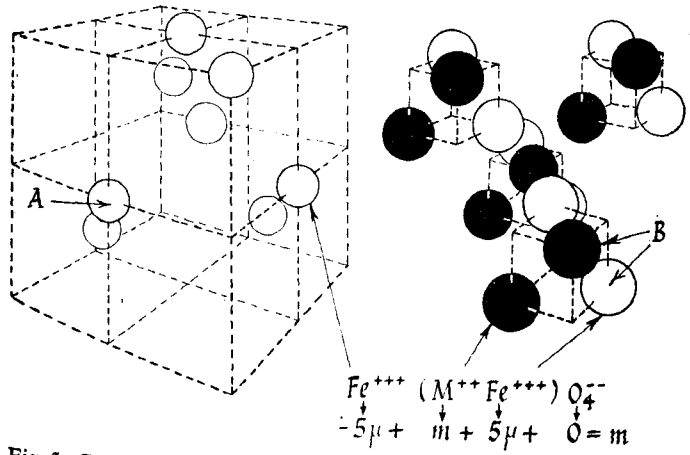


Fig. 5. Crystal Structure of an Inverse Spinel: for the explanation of the equation see p. 78.

right the white spheres also represent trivalent ions but on *B* sites, while the black spheres on *B* sites are the divalent ions.

The disposition of the trivalent and divalent ions shown in the figure corresponds to an inverse spinel structure to which all the magnetic ferrites belong. It is, however, possible to substitute for the divalent iron ions divalent ions of another suitable element. If, for example, the black spheres represented divalent manganese ions, Mn^{++}, the material would be a simple manganese ferrite. Similarly the element employed might be cobalt, nickel, or another divalent metal. The chemical formula for any simple magnetic ferrite might therefore be written $Fe^{+++}\ (M^{++}\ Fe^{+++})\ O_4^{--}$ where the symbol M^{++} stands for a divalent ion of any suitable element.

On the other hand certain ferrites may be formed having a normal spinel structure. The divalent ions then occupy the *A* sites, and the *B* sites contain only trivalent iron ions. Zinc and cadmium ferrites are of this type and these are not magnetic. For a zinc ferrite then, following the above system, we should write $Zn^{++}\ (Fe^{+++}\ Fe^{+++})\ O_4^{--}$.

Ferrimagnetism

The highest intensity of magnetization that can be uniformly achieved in one direction in a sample of iron is the same as the spontaneous saturation value of the domains. This occurs if the magnetic force applied externally is sufficient to cause the magnetization in all the domains to point in one direction. For pure iron the value of this bulk saturation magnetization is 21,580 gauss. In the case of magnetite, if all the magnetic ions contributed fully, a saturation value nearly as high would be obtained. Observation, however, shows that this is not so, magnetite having a magnetic saturation value less than one third of that of iron. An explanation of the comparatively low magnetic saturation values occurring in the case of all the ferrites has been put forward by Néel.[2]

It is first necessary to consider the magnetic moments of the ions occurring in the ferrites. A free oxygen atom has two electrons in the inner or K shell, two electrons in the 2*s* sub-shell, and four in the 2*p* sub-shell. The L shell of the free atom therefore contains a total of six electrons, which is two short of the eight required to complete it. In the oxygen ions in the ferrites these two vacant places are filled by electrons given up by the metal ions. Thus the oxygen ion O^{--} has its two main shells completely full. There are thus no uncompensated electron spins and no resultant magnetic moment, and so these ions make no direct contribution to the magnetism of the ferrite.

Figure 6, however, shows the conditions in other free atoms and positive ions. The 3*d* and 4*s* sub-shells are there shown for iron, manganese, nickel, and zinc. The divalent iron ion has given up the two 4*s* electrons and the ion has a magnetic moment of four units. The trivalent iron ion on the other hand has lost three electrons, two from the 4*s* sub-shell and one negative-spinning electron from the 3*d* sub-shell. The resultant moment of the Fe^{+++} ion is therefore five units. The free manganese atom on the other hand has a total of 25 extra-nuclear electrons, one less than iron, and it has, in fact, five electrons with positive spin in the 3*d* shell and none with a negative spin, so that the free

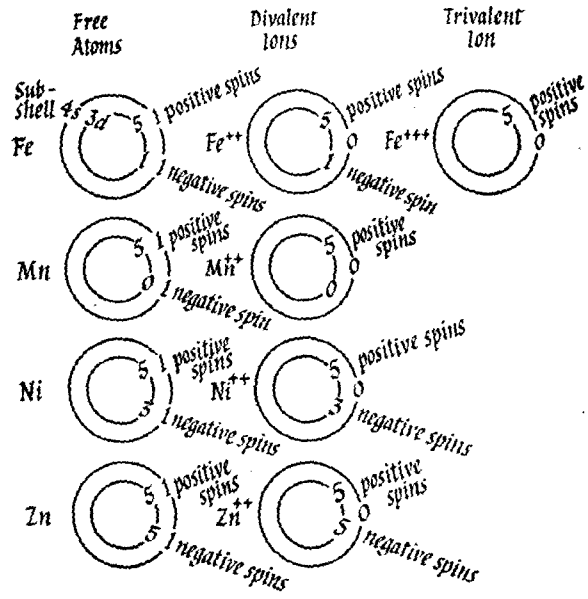


Fig. 6. The outer electrons in 3*d* and 4*s* sub-shells in each case. On the left are the free atoms and in the centre the divalent ions. The trivalent iron ion is also shown. The difference between the numbers of electrons with positive and negative spin, respectively, determines the magnetic moment.

manganese atom has a magnetic moment of five units. The di-alent manganese ion loses its two 4*s* electrons so that the Mn^{++} on still has a magnetic moment of five units. As a further example, the electrons in the outer sub-shells of nickel, which has vo electrons more than iron, are disposed in the free atom and e divalent ion as shown in Figure 6, so that the divalent nickel ns have a moment of two units. The figure also shows that alent zinc has zero magnetic moment.

Since oxygen ions are magnetically neutral, the magnetic pro-ties of the ferrites result from the positive ions present. These

ionic magnets situated on *A* and *B* sites are shown in Figure 5. It has already been mentioned that in metallic iron powerful quantum-mechanical forces of interaction exist between the spinning electrons responsible for the magnetic moments of the atoms, and that in ferromagnetism these positive forces lead to parallel alignment of the atomic magnets. It is, however, also possible for these strong forces to be negative, meaning that the neighbouring atomic magnets, while still on parallel axes, now point in opposite directions, just as in the case of an electron pair. According to Néel, in the ferrites, the interactions are in fact negative. This situation is further complicated by the presence in this instance of three sets of forces. There is the negative interaction between the ions on *A* sites. If this interaction predominated the result would be that half the ionic magnets on *A* sites, shown in the left-hand diagram of Figure 5, would point on parallel axes in one direction and the remaining half would be directed in exactly the opposite direction. There is also negative interaction between the ions on *B* sites, and once more a negative interaction between the *A* and *B* ions. It is considered that the *A*—*A* and *B*—*B* interactions are relatively weak, and are overwhelmed by the greater negative forces of interaction between the *A* and *B* ions. The final result of this is that all the *B* ionic magnets point in one direction and all those on *A* sites point the opposite way. Since in Figure 5 there are equal numbers of trivalent iron ions on *A* and *B* sites, respectively, the resultant magnetic contribution of these opposing ions is zero. The magnetic moment, or spontaneous magnetic saturation value of the ferrite domains is therefore the result of a parallel alignment of the *B*-site divalent ions shown black in Figure 5. Thus the magnetic moment per molecule would be *m* units, where *m* has the value five for manganese, four for iron, three for cobalt, and two for nickel.

The same argument applies also to ferrites that have a normal spinel type of structure. In zinc ferrite, for example, the divalent zinc ions occupy *A* sites as already mentioned, and the *B* sites contain trivalent iron ions only. Since the divalent zinc ions have zero magnetic moments, the *A*—*A* and *A*—*B* interactions are

also zero. Only the negative *B*—*B* interactions between equal trivalent iron ions remain, and there will now be no resultant magnetic effect. This phenomenon of equal and opposite arrays of ionic magnets is known as antiferromagnetism, while the preceding condition in which the arrays are unequal is said to be ferrimagnetic. Néel's theory is supported by Gorter's work.[3]

PRACTICAL ASPECTS OF FERRITES

The technologist is interested in the practical application of the ferrites in electrical equipment. The way in which the magnetic saturation value of the simple ferrites varies when the divalent element is changed has been described. Measurements on these materials, however, show that other magnetic and electrical characteristics also vary with composition. These include the magnetic change point, or Curie temperature, the magnetic directional properties in the crystals, the magnetostriction, the electrical conductivity, and the energy losses occurring as heat when the material is subjected to alternating magnetization. The Curie temperature, in this case, is that temperature at which the material ceases to be ferrimagnetic. This occurs when the internal thermal agitation has become so violent with rising temperature that the ordering effect of the forces of interaction has been overcome. Magnetostriction is the name given to the phenomenon whereby a magnetic material changes its dimensions when magnetized. Magnetite, for example, increases in length when magnetized whereas other ferrites contract. Theory and experiment indicate that the best magnetic properties are obtainable from materials of lowest magnetostriction.

It is not surprising then that if different ferrites are combined together wide variations in the magnetic characteristics can be obtained. The simple ferrites in fact form solid solutions one in the other. For example in Figures 3 and 5 half of the divalent ions shown by the black spheres might be of one element and half of another, or the proportions could be varied. Some of these mixed ferrites are found to have useful magnetic properties much superior to those of simple ferrites. Basic knowledge of this subject is, however, not sufficient to enable the

characteristics of any composition to be predicted, and hence much experimental work of an empirical or semi-empirical nature is proceeding.

Ferrites are finding applications in communication circuits, in components used in radio and television receivers, including the use of built-in ferrite-rod aerials, and in data-storage elements in electronic computers. There are also applications of a highly technical and specialized nature[4] in equipment transmitting and receiving electromagnetic radiations in the microwave region, that is, at frequencies of the order of 10,000 megacycles per second. The ferrites, because they can be made with such very low electrical conductivities, are the only magnetic materials available for most of these applications. Thus the lodestone, scientifically developed, once more comes into its own.

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MALE OR FEMALE?

D. B. CARLISLE

MANY species of animals show a sexual dimorphism; that is to say, the two sexes are externally different. The differences between a man and a woman are obvious, but in geese there appear to be no external differences. Even in geese, however, the internal anatomy is different in the two sexes; but in many of the 'lower' forms of life the only differences we can observe are in the gonads themselves – that is to say, the ovary is different from the testis. Here all other parts of the anatomy may be identical in male and female, so far as we can see. An example of an animal of this kind is the cockle. We say that cockles differ only in the primary sexual characters–the gonads; geese differ also in the accessory sexual characters – those organs and ducts that are directly concerned in the reproductive act or in mating; while men differ from women also in secondary sexual characters – there are differences between the sexes that have nothing directly to do with reproduction – differences in hair pattern, timbre of voice, and so on. It is only when there are such secondary sexual differences that we refer to a sexual dimorphism; because, of course, all animals in which there are separate sexes have at least primary, and often accessory, sexual differences.

CONTROL OF SEX DIMORPHISM

Strictly speaking, the sex of an animal is defined by the functional condition of the gonads and of the whole reproductive system: an animal that is *functioning* as a male is a male; one *functioning* as a female is a female. The male function involves, as a minimum, the production and release of sperm; the female, the production and release of eggs. An animal intermittently performing these functions is assigned to the corresponding sex.

Geneticists commonly define sex in terms of their own discipline, but I believe that this may become misleading. If sex is defined in genetical terms, then it is impossible to talk about sex reversal, for the genetic constitution of an animal remains constant: there is no reversal here. Such a reversal is in the functional sex of an animal. Where the genetic constitution of an animal is in question then the term 'genetical sex' should be used, or some equivalent expression.

It is only in rare individuals that there is a failure to make the accessory and secondary sexual characters correspond with the primary sex, the sex of the gonad. In medical parlance such individuals are termed pseudohermaphrodites and, for the moment, nothing more will be said about them. How, then, is this almost unfailing correspondence between primary sexual characters and secondary and accessory characters brought about? How does it happen that a man possesses a penis, a deep voice, and a beard, to go with the testes, or a male stag beetle possesses the antlers and copulatory appendages?

Two possible mechanisms by which this may be attained are well known. By the first, the most direct mechanism, each cell in the body takes on the sexual characters appropriate to its position in the body, purely because of its own genetic constitution, without reference to any other cell or tissue. If the process of cell division takes place normally during development, every cell possesses exactly the same complement of genes as every other cell in the body and all are, therefore, of the same sex. The function and morphology of a cell are determined by its position in the body. Once these are determined, then, the cell, by virtue of its own genes, takes on the sexual characters appropriate to it. If, experimentally, we graft a wing rudiment from a male to female moth, in a species in which there is a sexual difference in pattern, the wing develops with characteristic male pattern, corresponding to the genetic sex of the constituent cells. The process of cell division is not, however, without its accidents, and the chromosomes may be unequally divided between the daughter cells. This will result in the genes that determine sex being in a state of imbalance so that the sexual characters of

the patch of tissue that results from the repeated division of one such cell may show aberrations. Worse still, if the 'sex chromosomes' are involved in such an unequal division, one of the daughter cells may be of the opposite genetic sex, and hence the patch of tissue resulting will be of the opposite apparent sex to the rest of the body. If such an unequal division takes place late in development the resulting inappropriate patch is, of course, small, but if it takes place early, a large part of the animal may show the characters of the wrong sex, or the two halves, even, may be of opposite sexes. Such an animal, with a patch of tissue inappropriate to its sex, is termed a gynandromorph. If one

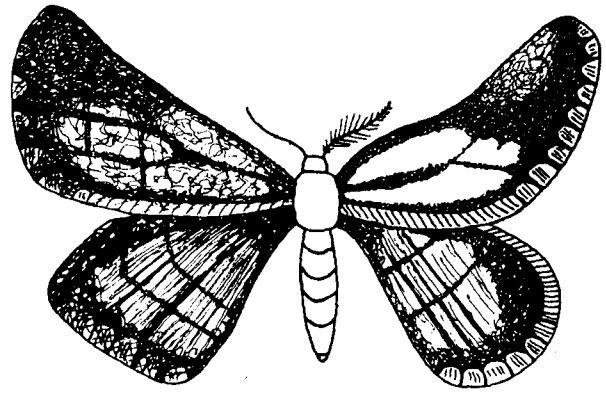

Fig. 1. A gynandromorphic pine moth, in which the left half is female and the right half male.

gonad is affected, then it may become fully functional in the opposite sex and we have a true hermaphrodite, with functional ovary and testis. This type of sex determination is especially characteristic of insects.

A totally different mechanism is found in birds and mammals, and probably other vertebrates. Here, by the production of hormones, the gonads exert a controlling influence on the secondary and accessory sexual characters. The genetic constitution makes itself apparent upon sexual characters only within the

gonads. These then secrete into the blood various hormones – chemical substances – which stimulate other tissues of the body to take up the characters appropriate to the sex of the developing animal. If we perform on birds the experiment of grafting a feather rudiment from a hen to a cock, the feather does not develop into a hen feather, but into a typical cock feather, quite unlike the situation we found after grafting the wing rudiment of a moth. The tissues of either sex retain the ability to respond to the hormones secreted by the gonads, and a buck rabbit can be made to yield milk by suitable injections of female hormones.

In the course of development of the gonads in birds and mammals a pair of organs each capable of both male and female reproductive activities appears, with two parts, one secreting male and the other female hormones. After a short time, the part that has the genetic ascendancy, the part, that is, whose sex corresponds to the genetic sex of the cells, secretes more and more of its hormones, which inhibit the other part so that it atrophies more or less completely. From then the sex is determined and the secondary and accessory characters follow suit, stimulated by the hormones secreted by the developing gonad. By experimental interference it is possible to inhibit the part of the gonad corresponding to the genetic sex of the animal, taken early enough, and allow the other part to obtain an ascendancy; once this has happened it will not normally lose its ascendancy so that the final sex of the animal will be opposite to the genetic sex. This sort of thing can also happen in nature, should some accident or disease early in antenatal life interfere with the developing gonad. In the hen it can happen even later, for here the ovary develops only on one side. On that side it completely suppresses the rudiment of the testis, but on the opposite side the testis rudiment persists. If now the ovary is destroyed by disease (or experimentally) leaving the opposite testis rudiment intact, then this can develop, once the inhibiting influence of the ovary is removed, until it becomes a fully functional testis. The secondary and accessory sexual characters follow suit and the erstwhile hen becomes a cock capable of fathering chicks. This is true sex reversal brought about by pathological means.

Such a mechanism of controlling the secondary and accessory sexual characters is open to a different kind of accident from that found in insects, where gynandromorphs are the most common misfits. If the gonads secrete either excess or insufficient hormone there may be corresponding over or under development of the accessory organs. Or one or more organs may be excessively sensitive to either male or female hormone – and each individual secretes a little of the hormones of the opposite sex – so that a man may develop an excessive growth of hair, for instance, if his beard is over-sensitive to the male hormones, or well-rounded breasts, if these organs are excessively sensitive to the female hormones produced by his testes. Conversely, certain organs or the whole body may be insensitive to either type of hormone. If this is combined with excessive sensitivity to the hormones of the opposite sex, then the result is the pseudohermaphrodite, with gonads of one sex and most or all of the secondary and accessory sexual organs corresponding to the opposite sex. The true hermaphrodite, possessing both ovary and testis, is rare in birds and mammals, and, so far as I know, never functional either as a male or as a female. This condition can arise by the same sort of accident of cell division that produces gynandromorphs in insects, but only when this affects the cells that go to form the gonads. Then one gonad develops as a testis, while the other becomes an ovary. Since, however, they are both secreting their hormones, which tend to inhibit the other type of gonad, each is inhibiting the other and neither becomes fully functional. Contrast this with the condition in insects, where a true hermaphrodite, developed by gynandromorphy, may be fully functional in both sexes, since neither gonad inhibits the other.

In vertebrates the pituitary gland, situated at the base of the brain, exerts a controlling function over the gonads, but even this has no direct bearing on sex. The gonadotrophic hormones secreted by the pituitary gland stimulate equally both testis and ovary, and it is only when the hormones from the gonads have, in their turn, stimulated the pituitary that it produces, in the female, the special hormones concerned with egg-laying,

mother-care, or pregnancy. Despite the control exerted by the pituitary over the gonads, it is they, not the pituitary, that have the final control over the secondary and accessory sexual characters. The pituitary, in this context, serves primarily to regulate the breeding cycle once the sexual form of the animal has been determined by the gonad.

Such, in outline, are the classical pictures of the two mechanisms whereby the secondary and accessory sexual characters are made to correspond with the sex of the animal. In insects each and every cell of the body takes up the sexual characters appropriate to its own genetic constitution. In birds and mammals only the gonads do this and they, by means of hormonal secretions into the blood, stimulate the other tissues and organs of the body to take up the appropriate sexual characters, and in the development of the body both types of gonad are formed at first, till one overcomes the other and prevents its further development. It is obvious that this type of control of sexual dimorphism is incompatible with true, functional hermaphroditism, whereby a single individual can have functional gonads of both sexes, both an ovary and a testis.

SEX REVERSAL

So far we have confined our attention to the best-known mechanisms of control of sexual dimorphism. A few years ago it would have been possible to regard these as the only known mechanisms, but of recent years investigations have revealed a little of several other mechanisms for attaining this end. Almost all the higher crustaceans, for instance, show a sexual dimorphism, for it is the female's habit to carry the fertilized eggs about with her until they hatch. This entails modifications of anatomy and physiology. Moreover, since fertilization of the egg is internal there is an act of copulation, and this involves the development of external accessory sexual organs concerned with the act of mating. In addition there may also be secondary sexual differences in many species, particularly in those species in which the males fight for the females; thus many male crabs have larger claws than the females. Even in prawns, a mere

glance is sufficient, after practice, to reveal the sex of an animal, such is the sexual dimorphism.

Prawns

Now in a few species of prawns a peculiar type of hermaphroditism is found, whereby each individual functions for a time as a male and then reverses sex completely and irrevocably, in all primary, secondary, and accessory sexual characters, to become a female for the rest of its life. This occurs only in a very few species, which are widely scattered and unrelated. Within the genus *Pandalus*, *P. borealis* – the large, northern, deep-sea, red prawn – reverses sex in this manner; and so does the Japanese species *P. kessleri*; while the British species *P. bonnieri* is, like most other prawns, an animal with two separate sexes – the male never reverses sex to become a female. The unrelated Mediterranean prawn *Lysmata seticaudata* and the blind burrowing prawn *Calocaris macandreae* are other species in which sex reversal takes place, so that all the males that survive long enough become females. *Hippolyte*, which is closely related to *Lysmata*, has, like *Pandalus bonnieri*, two quite separate sexes, and so do almost all other prawns and shrimps.

For this type of hermaphroditism to turn up so sporadically amongst prawns, its control must be by some quite minor variation of the normal mechanism of control of sexual dimorphism within the group as a whole. Such a type of hermaphroditism could not arise under either of the well-known types of control of sexual dimorphism I have outlined. The control of sexual dimorphism in these hermaphrodites, therefore, must be by some third mechanism and we must presume that its control in prawns in general must be different also.

Let us consider the sexual biology of one of these hermaphrodite prawns, *Lysmata seticaudata*. The eggs hatch in about May and the young animals then spend about three months free-floating in the sea several hundred feet deep. About August they move inshore and settle down to life on the bottom in a few feet of water. From this stage onwards the gonads of these juveniles may be seen developing. From the start testis and

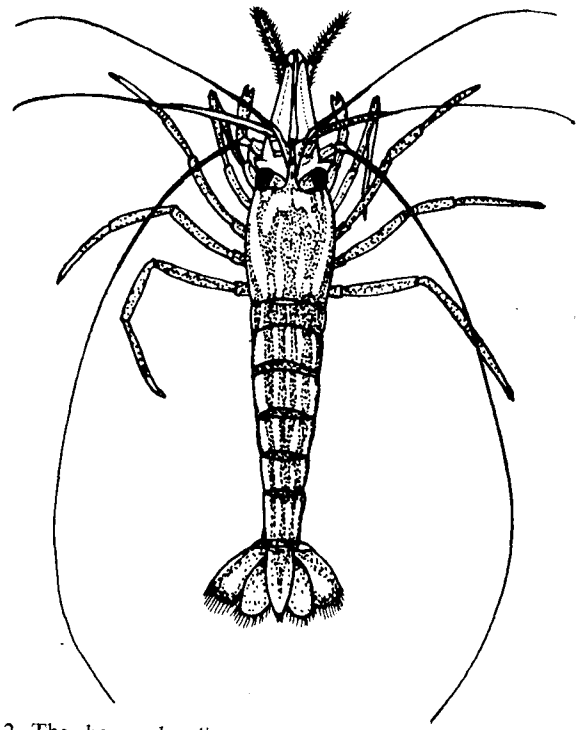

Fig. 2. The hermaphrodite prawn, *Lysmata seticaudata,* which spends a year as a male and then reverses sex to become a female for the rest of its life. The pattern on its back is made up of thousands of red dots, in groups separated by white lines.

ovary develop side by side, neither inhibiting the other. About January, five months after settling down to life on the sea-bed, *Lysmata* attains sexual maturity as a male. The ovary ceases to grow, while the testis has developed to full activity, and the animal can now copulate as a male. During the period of sexual immaturity the male accessory and secondary sexual characters have been appearing, and coincident with maturity of the testis

the full male habit and appearance are attained. During the following year or eighteen months the male *Lysmata* grows, like all crustaceans, by casting off the old shell and developing a new one a size larger; and each new shell bears the male external characters. Then after about a year, or a little longer, of male life, while the testis is still fully functional, the ovary begins to grow again. The next time the shell is cast the new one no longer bears the male appendages. It may be sexless in appearance, having neither the characters of a male, nor those of a female; but in quite a large proportion of individuals the new shell emerging is that of a fully fledged female, so that I have observed a single individual copulate as a male, cast its shell 90 minutes later, and within five minutes copulate as a female. Once the external appearance is that of a female, the male ducts are blocked, no longer opening to the outside, and, as in a tom-cat in which the veterinary surgeon has ligated the male ducts, the testes degenerate, until after about another year or so they are a mere remnant, and a year later they have disappeared. Two and a half years as a female seems to be about the utmost limit of its life-span, during the whole of which time the animal is growing with repeated castings of the outgrown shell. In other known species of hermaphrodite prawns, the sexual biology, though different in detail, is broadly the same.

It is unfortunate that almost nothing is known of the early development of the gonads of prawns; we do not know whether or not they develop like those of birds and mammals, with first a bisexual gonad of which one part later disappears. It seems apparent, however, that in the hermaphrodite prawns neither the ovarian part nor the testicular part of the gonad is inhibiting the other. Nor in more normal prawns, with separate sexes, have extracts of testes been shown, when injected into females, to exert an inhibitory effect on the ovaries, neither have ovarian extracts been found to inhibit the testes – two events that follow such injections in mammals and birds. It seems then that in both hermaphrodite and normal prawns neither gonad is secreting hormones that tend to inhibit the other. Moreover, testis extracts have never been found to exert any action in stimulating the

development of the male secondary and accessory sexual characters, and extracts of ovaries only stimulate the development of the special organs for carrying the eggs, which appear cyclically – the pregnancy organs we may call them. Conversely, by destroying the ovary it has been shown that prawns lacking this organ develop the normal female characters, but not the pregnancy organs. It is, therefore, evident that the only sex hormone produced by the gonad is a pregnancy hormone: the main secondary and accessory sexual characters are not controlled by hormones secreted by the gonads. The control of dimorphism is thus not the same as that found in birds and mammals. Nor, probably, is it like that in insects, for in many tens of thousands of prawns I have never seen any trace of gynandromorphism. Moreover, if sexual dimorphism were directly controlled genetically in *Lysmata,* as it is in insects, then the genes concerned in the secondary and accessory sexual characters would have to be rate genes, or delayed action genes – genes, that is, controlling the rate of development of a character or controlling the moment of appearance of a character. Now such genes are notably, but unequally, influenced by temperature. Alteration of the temperature of laboratory stocks would alter the rate of action of such genes and produce aberrations, produce, that is, animals with some of the characters developing out of turn. In *Lysmata* no such thing happens: whatever the temperature, all changes take place together. This suggests that there is some overriding control, probably hormonal, of all the changes associated with sex-reversal, and we must look for this in some slight modification of a mechanism for controlling sexual dimorphism in more normal, bisexual prawns.

Some ten or a dozen years ago a French worker, J. B. Panouse, discovered in prawns the existence of a hormone that prevented the development of the ovary. For most of the year this is released into the blood at a steady rate, so that the ovary is very small. Then, as the breeding season approaches, its secretion stops and the ovary is allowed to increase in size and to mature till it is ready for egg-laying. After the eggs have been fertilized

and laid, the secretion of the hormone begins once more so that the exhausted ovaries cannot start preparing for another breeding season prematurely. This ovarian-inhibiting hormone, as it is called, is produced by a complicated organ, at the back of the eye, called the *X* organ-sinus gland complex. The *X* organ-sinus gland complex bears a great resemblance in many ways to the pituitary gland of vertebrates and secretes a number of hormones which exert some control over such phenomena as diabetes, colour change, growth, casting of the shell, and rate of urine flow – just like the pituitary in fact, except that vertebrates do not need to cast a shell. If the *X* organ-sinus gland complex is removed surgically the ovary can begin to develop and increase in size at any season of the year. Within a month it may be 70 times as big as before the operation.

It is noteworthy that male prawns have at all seasons of the year quite large amounts of the ovarian-inhibiting hormone, but, unlike the gonadotrophic hormone of the pituitary of vertebrates, which affects equally both ovary and testis, this hormone has no effect whatsoever on the testis of prawns. I have searched for evidence of a hormone that might inhibit the growth of the testis but failed to find anything that might indicate the existence of one. The only happening that causes degeneration and inhibition of the testis is closure of the male ducts, which in all animals, including man, invariably leads to degeneration of these organs. Recently Japanese workers have suggested that in female prawns there is another hormone that promotes the development of the ovary; this hormone seems to be absent from the males. Once more it is a product of the complex system of glands at the back of the eye. The ovarian-inhibiting and the ovarian-stimulating hormones seem to act in opposition to one another on the ovary; nothing is known of the action of the latter hormone on the secondary and accessory sexual characters. On the other hand the ovarian-inhibiting hormone is certainly responsible for preventing the premature appearance of feminine characters in immature females, and for preventing their appearance in males. There is some evidence that if male prawns are parasitized, the *X* organ-sinus gland complex ceases

to produce the ovarian-inhibiting hormone. This does not lead to the development of an ovary, but it does lead to a feminization of all or most of the secondary and accessory sexual organs.

We are beginning to obtain a picture of the control of sexual characters by hormones secreted by the glands at the back of the eye. But this is not the whole story. About three years ago Manfred Gabe, working in France, discovered in crustaceans another organ which has been shown to secrete hormones which affect the gonads and the sexual characters. This organ, situated in the head, he called the *Y* organ. It has been shown that if this organ is removed from juvenile crabs – the experiment has not yet been performed upon prawns – the gonads do not develop any further towards maturity, but rather degenerate, while there is no progress towards attaining the adult sexual dimorphic characters. Once maturity has been reached, however, removal of the gland does not interfere in any way with the reproductive organs or with the secondary and accessory sexual characters.

If we put together these various pieces of information, with some others not mentioned here, the picture that is presented of the sexual development of prawns and crabs is something like this. During childhood the *Y* organ is secreting a hormone which stimulates the growth of the reproductive organs and the characters that go with them. This may well be an indifferent stimulus, stimulating either testis or ovary to develop. The primary determination of sex is likely to be in the glands at the back of the eye. In the female these produce a hormone promoting the development of an ovary and of female characters. This is counterbalanced to some extent by the ovarian-inhibiting hormone, secreted by the *X* organ-sinus gland complex, which, by restraining the development of the ovary and the sexual organs, prevents premature attainment of sexual maturity. In the male the hormone promoting the development of an ovary is absent while more of the ovarian-inhibiting hormone is produced, completely suppressing any tendency to produce an ovary or female sexual characters. There seems little doubt that there must also be some organ somewhere in the body that is promoting the development of a testis and the male sexual organs, but this has

not yet been located. However this may be, it seems certain that the high production of the ovarian-inhibiting hormone is necessary to prevent the appearance of feminine characters in the male. Crabs show one difference from prawns: the glands behind the eye produce a hormone that inhibits the ovary, but this seems only to come into action after the first time a crab has bred. A male prawn is ready for breeding throughout the year; a male crab is only ready to mate during the breeding season.

With sexual development controlled in this way, how can the hermaphrodite species arise and develop? Some idea can be obtained by imagining a male prawn that for some reason is secreting the ovarian-promoting hormone. Since the testis does not produce any hormones that inhibit the ovary, and the ovarian-promoting hormone can apparently cause the production of an ovary even in the face of the opposition produced by its opponent, then this male prawn will come to possess a rudimentary ovary. If, now, the secretion of the ovarian-inhibiting hormone ceases, immediately feminization of all the characters begins and the ovary can develop to full maturity. Once the exits of the male ducts have been closed by feminization of the shell of the animal, the testes inevitably degenerate. Experiments I performed in 1956 on *P. borealis* confirm this suggestion.

We have in prawns, then, a third, quite distinct, mechanism for the control of sexual dimorphism; for the regulation, that is, whereby a male possesses the male secondary and accessory sexual organs, and a female the female organs; a method which neither makes it necessary that every cell should determine its own sex (as in insects) nor that the gonads should control the sexual characters of the rest of the body (as in birds and mammals). Rather, the sex of the gonads themselves is controlled by some overriding sex-determining complex of glands, which regulates all the sexual characters, primary as well as secondary and accessory. Indeed it has been shown that in sand-hoppers (sometimes called sand-fleas, although they are not true fleas, nor even insects, but crustaceans), if an ovary is grafted into a male it actually becomes converted into a testis, even if the male is first castrated. If the same operation were to be performed in a

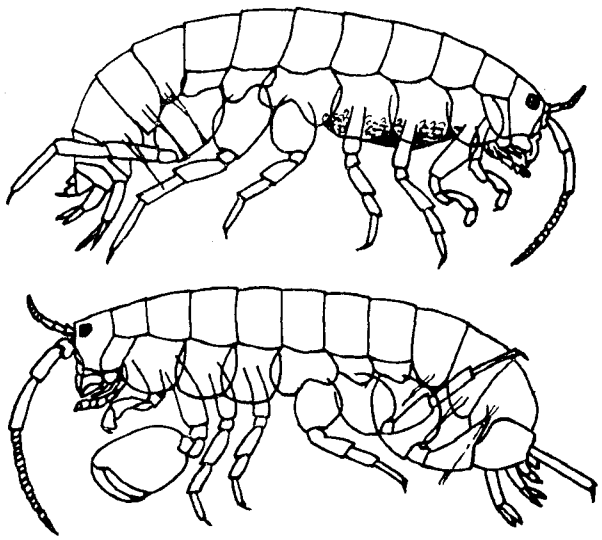

Fig. 3. The common sand-hopper of British shores. The upper drawing shows a female and the lower a male. Notice the differences, especially the egg-pouch between the front legs of the female, the large claw of the male, and the differences between the big, hind, jumping legs.

mammal or bird the result would be feminization of the body, not masculinization of the ovary, while in an insect neither body nor gonad would change. The gland that is responsible for this effect on the ingrafted ovary is a tiny triangular gland which lies in the wall of the male duct, and it seems to be this gland that in the first place stimulates the production of a testis in the male sand-hopper. A gland of the same kind exists in wood-lice (sow-bugs) also, but here it is embedded in the tissue of the testis. Nothing has been shown to exist in these animals that exerts the contrary influence, whereby a testis grafted into a female might become converted into an ovary.

A gland stimulating the development of the testis may possibly exist also in prawns, but if so it must be incapable of acting

on ovarian tissue, only on predetermined testicular tissue, for otherwise the type of hermaphroditism exemplified by *Lysmata* could not exist.

It is apparent that at least five hormones are concerned in the regulation of sexual dimorphism in crustaceans, but we know nothing about the chemistry of four of them. The ovarian-inhibiting hormone, which has been known longest, is the exception. Recent evidence suggests that this, like the sex hormones of vertebrates, is one of the group of compounds known as steroids. It can be injected into prawns, where it will cause the abrupt cessation of all growth of the ovaries, or it can be fed to them, when rather larger amounts are needed to produce the same result. It is noteworthy that queen bees produce a similar substance, apparently also a steroid, which is secreted over the surface of the body and is thence licked up avidly by the worker

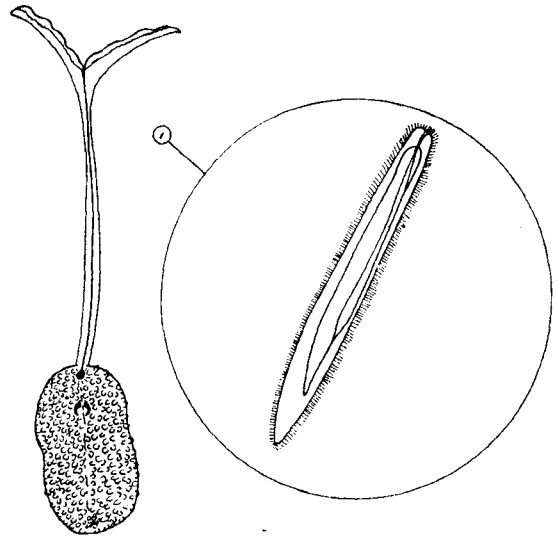

Fig. 4. The worm *Bonellia*. To the left is a female, and in the centre is a male to the same scale; on the right is shown the male much enlarged.

bees. So long as they are receiving an adequate supply of this substance their ovaries are completely inhibited – worker bees are, of course, sterile females. But if the supply fails for any reason, their ovaries begin to grow until some of them can lay eggs. If they are fed with the substance obtained by washing queen bees in alcohol, this is prevented. Feeding them on the ovarian-inhibiting hormone of prawns has the same effect. Conversely, if the queen bee substance is injected into prawns, their ovaries are inhibited in the same way as they are by the hormone.

Recent evidence suggests that in the worm *Bonellia* steroids also play a part in regulating sexual characters. In this worm the female is several thousand times as large as the male, which is an internal parasite in the excretory organs of the female. After hatching from the egg the young worm swims around feeding for a short time. If during this period it comes into contact with the long proboscis of a female *Bonellia* it remains there and becomes for a time firmly fixed. It has been shown that the proboscis of the female secretes some substance, apparently a steroid, which prevents the young adherent worm from developing ovaries; instead, the young one develops testes, without ever changing to the adult worm shape. Later this male, which never grows any larger, moves down the groove in the proboscis, enters through the mouth of the female (near the base of the proboscis) and settles down in the excretory organs. A young worm which never finds the proboscis of a female to settle on, continues growing, settles on the sea bed and changes its shape to become the adult female.

Can it be that in insects, vertebrates, worms, and crustaceans, with quite different mechanisms for the control of sexual dimorphism and reproduction, steroids are always concerned in some way or other in regulating sexual development? What is the fundamental biochemical similarity between them? I am sure there must be one.

ULTRASONICS IN INDUSTRY

A. LINFORD

To an ever greater degree, the natural properties of materials are being utilized to provide tools for industrial processes, and expansion in this direction has been vastly increased by advances in electronics which make it possible to detect, amplify, and measure the instantaneous value of minute electrical quantities fluctuating at great speed.

One recent development has been the application of mechanical vibration to a wide variety of engineering problems ranging from thickness measurement to drilling square holes. Very high frequencies are used, well above the normal audible range. Among the advantages of using ultrasonic vibrations is that these vibrations can readily be 'beamed' in the required direction, and a very short wavelength is of paramount importance in thickness measurement, and other similar applications.

The fundamental laws governing mechanical vibration, whether audible or ultrasonic, are quite simple. The velocity of the wave is equal to its frequency multiplied by its wavelength; the velocity of the wave is determined by the medium, e.g., steel, copper, or water, through which it is travelling, and the waves are reflected back along the path of approach when they strike an interface separating two materials having different acoustic properties. The velocity of sound in any material is constant and it is not affected by the form of treatment the material has undergone. For example, in steel, the velocity of sound is the same whether the steel has been heat treated or case hardened. In addition, when the distance of the reflecting surface from the surface at which the vibrations are generated is equal to half the wavelength, or any whole number multiple of this length, i.e., to any harmonic of the fundamental wavelength, the reflected wave will reinforce the incident wave, and resonance will

occur. Resonance is obtainable when the distance involved is equal to half a wavelength or a whole number multiple of this because the maximum amplitude of oscillation occurs twice per wavelength. An analogy is an organ pipe which will resonate when the length of the air column within the pipe is equal to half the wavelength generated or to any harmonic of this.

Production of Ultrasonic Vibrations

Before thought can be given to the application of the phenomenon, some means are necessary for producing it. The simplest method is, of course, to utilize the principle of the whistle or siren. While such systems have certain applications, their scope is limited as mechanical difficulties prevent them from generating vibrations of frequencies much above the sonic range.

For ultrasonic vibrations magnetostrictive, piezo-electric, and electrostrictive vibration generators are used. These convert electrical energy into mechanical energy and vice versa. The operation of magnetostrictive generators depends on the fact that when they are magnetized forces are set up in magnetic materials and these forces cause an alteration in the dimensions of the magnetic materials. If, therefore, an alternating current is passed through a coil surrounding a rod of magnetic material, its length will change in sympathy with the frequency of the alternating current. If this frequency coincides with the natural frequency of the rod, the latter will resonate.

The frequency range of a magnetostrictive transducer is limited to a maximum of about 100,000 cycles per second, so for higher frequencies piezo-electric and electrostrictive devices must be used. The piezo-electric effect occurs in certain artificial and natural crystals, notably quartz. The thickness of such crystals alters when a difference of electrical potential is applied across opposite faces. Here again, resonance is obtained when the applied frequency of the electric supply corresponds to the natural frequency of the crystal.

Electrostrictive transducers usually consist of barium titanate crystals which exhibit the same characteristics as quartz crystals, with the difference that they will vibrate in any direction. Both

piezo-electric and electrostrictive transducers can be used for frequencies up to millions of cycles per second.

Thickness Measurement

One of the most important applications of ultrasonics is to the measurement of the thickness of metals, alloys, and a wide range of other materials when access can be had to one side only. During manufacturing processes, conventional forms of measurement, e.g., micrometers and similar measuring devices, are available, but it is often a difficult and lengthy task to get the desired dimension when only one of the two surfaces is readily accessible. A case in point is checking of the thickness of the wall of a pipe on the outside of a bend, in which region the thickness of metal may be considerably reduced. Metal spinnings of complicated shape also present difficulty in mechanical gauging. Even more important is the problem of thickness measurement in instances where one wall is quite inaccessible. For example, in the manufacture of lead-covered cable, if the thickness of the lead sheath is to be measured, measurement must be carried out from the outer surface only because the inner surface can only be reached by destroying the cable.

In the chemical, petroleum, and gas industries particularly, periodic checks of the wall thickness of pipes, fractionating columns, pressure vessels, and so on are desirable as an insurance against the possibility of dangerous conditions resulting from corrosion, and also to obtain optimum useful life from the equipment. Unless some means for carrying out the necessary measurement from one surface only are adopted, the equipment to be checked must be shut down specially – an expensive and lengthy procedure if, for example, a long oil pipeline is involved. But ultrasonics make it possible to measure thickness without in any way interfering with the operation of the plant – measurements can be made while the plant is under pressure or at an elevated temperature or both. In ship repairs a particularly large economy is achieved by ultrasonic thickness measurement. In the past it has, for example, been necessary to resort to drilling holes in the plates in order to check the degree of

corrosion, and after measurement the test holes had to be plugged: in a tanker upwards of 1,000 holes may have to be drilled and plugged. (Inset 19.)

While such methods as back-scattering of radioactive sources may be used, the application of these is limited. For example, β-ray gauges can only be used for very thin metal sheets and precalibration against a metal sheet of known thickness is required, i.e., the method is a comparative one. Sources of γ-rays are more penetrating, but they are of high energy, and so may be dangerous to personnel in the vicinity unless heavy shielding and adequate safety precautions are adopted. While the ultrasonic meter also has certain limitations, it does not suffer from the disadvantages just mentioned.

From a consideration of the laws of mechanical vibrations, it is apparent that for thickness measurement an echo method could be used, as in the marine 'Asdic', in which the time for a pulse to be transmitted through the medium (i.e., water) and back again after reflection from the far surface is measured. Knowing the velocity of sound in the medium, the distance can be found by simple arithmetical calculation. Unfortunately, the time intervals involved are minute and a cathode ray oscilloscope is required to present against a time scale the instant of departure and return of the impulses. A cathode ray tube is relatively bulky and fragile, and is not conveniently operated from batteries. Equipment of this nature, therefore, is transportable rather than portable, and can only be used where a power supply is available. In addition, when dealing with small thicknesses, below half an inch or so, this pulse method tends to be inaccurate. The trace produced on the oscilloscope screen is difficult to interpret and, to get consistent readings, a high degree of skill is required.

The alternative system of measurement employs resonance. The virtue of this method is that it requires no measurement of time, so that the measuring instrument can be simple and robust. In addition, if desired, the equipment can be operated from dry batteries. Consequently it is self-contained and portable – it is small and light so that it can be carried by the operator. This

enables thickness measurements to be made accurately and rapidly in locations inaccessible to bulky equipment.

In the resonance method of thickness measurement, the frequency of the vibration applied to the workpiece is gradually increased, i.e., the wavelength is decreased, until resonance is obtained. If this is the fundamental wavelength, knowing the velocity of sound in the material under test, the thickness can readily be calculated, since the thickness will be exactly equal to half the wavelength. It may be, however, that for the thickness of the material, the fundamental wavelength does not lie within the range of the instrument, or it may be missed and one of the harmonic resonances obtained instead. The method employed, therefore, is to note the frequency at one resonant point and then gradually increase the frequency until the next resonant point is reached. It may be shown theoretically by an application of the laws of mechanical vibration already quoted, that the difference in frequency between two adjacent resonant points, i.e., the difference between the 2nd and 3rd harmonic frequencies, or between the fundamental and 2nd harmonic frequencies is equal to the fundamental frequency. Consequently, the thickness measured is equal to the velocity of sound in the material divided by twice the difference between the frequencies at which resonance is obtained. The difference must be doubled because the thickness is equal to, or is a whole number multiple of, *half* the fundamental wavelength.

The importance of applying frequencies far above the audible range can be illustrated by a simple example. It may be assumed that the problem is to measure the thickness of an aluminium wall, in which the velocity of sound is 20,600 feet per second. Assuming the limit of the audible range to be 20,000 cycles per second, the wavelength produced (velocity divided by frequency) is roughly one foot, so that the *minimum* thickness that could be measured is in the region of six inches. In practice, with commercial equipment, frequencies in the region of two million cycles per second can be produced, which enables thicknesses of the order of 1/16 in. to be measured accurately.

For converting electrical energy into mechanical energy a

piezo-electric generator is generally used. This consists of a quartz crystal cut into the shape of an 'X' and ground to resonate at the desired frequency. On account of practical considerations, the frequency range is generally limited to a range of 2 to 1. For example, a frequency range of 0.75 to 2 million cycles per second may be selected. If only fundamental frequency resonance were employed, the range of thickness that could be measured would be of the same order as the frequency range, but by using harmonics in the manner already described, the thickness range can be vastly extended, e.g., it may be of the order of 100 to 1. For example, in a sheet of aluminium 0.1 in. thick, the fundamental frequency equals the velocity of sound in aluminium (20,600′ per sec.) divided by twice the thickness in feet, i.e., divided by $2 \times 0.1/12$ which is approximately 1.24 million cycles per second; this is within the instrument range of 0.75 to 2 million cycles per second. On the other hand, for a block of aluminium 10 in. thick, the fundamental frequency is 100 times less, i.e., 12,400 cycles per second, but the harmonics from 70 (0.868 million cycles per second) to 150 (1.86 million cycles per second) all fall within the frequency band, and any two adjacent harmonics can be used to obtain the fundamental frequency and hence the thickness.

The alternating electric supply applied to the transducer is derived from a variable frequency electronic oscillator, and the output frequency of the oscillator can be swept over the full range by turning a knob over a graduated scale. When the oscillator is tuned to an integral multiple of the fundamental wavelength in the material under test, there will be an increase in the combined signal, since the outgoing signal from the transducer will be reinforced by the reflected (incoming) one. This will cause pulses of current in the anode circuit of the electronic valve; the oscillator is tuned about the thickness resonance frequency at a predetermined number of times per second within the audible frequency band. These electric pulses are amplified and either indicated on a dial or applied to a microphone. All the operator has to do is to rotate a knob until the sound produced reaches its maximum value, note the dial reading and continue rotating until the sound reaches a second maximum, when the dial reading is again noted. By subtracting the two readings, the fundamental frequency is obtained. The instrument may be provided with a built-in computer. The point of maximum sound intensity can be estimated with a high degree of accuracy, the guaranteed accuracy of commercial equipment used under industrial conditions being of the order of ± 3 per cent, but accuracies of ± 1 per cent can often be obtained in practice.

An alternative arrangement which enables direct readings on a scale to be obtained, e.g., calibrations in steps of one thousandths of an inch, entails the use of a cathode ray oscilloscope. An instrument of this type is particularly suited for repetitive work, since it entails no calculation and is self tuning, i.e., thickness indication is automatic. The oscilloscope will display harmonic resonance as well as the fundamental and therefore the frequency range must not be greater than 2 to 1. To increase the scope of the instrument, its total frequency range is split up into a number of overlapping bands, the frequency ranges of which do not exceed 2 to 1, and a separate transducer used with each. For example, with a total frequency range of 0.75 to 8 million cycles per second, six overlapping bands are used. This enables the thickness resonance to be displayed for materials varying in thickness from about 0.015 in. to 6 in. Here it may be stressed that the exact maximum and minimum thickness that can be measured by a particular instrument is determined by the velocity of sound in the material under test.

To carry out a measurement, the crystal, of the order of 1 in. diameter, is mounted in a holder and firmly pressed against the surface of the material. To facilitate the work of the operator, when dealing with ferrous metals, the transducer can be held in position magnetically. To prevent excessive wear of the crystal, which would be occasioned by its being moved continually over various surfaces, its face is usually protected, e.g., by silvering or by chromium plating. For particularly arduous conditions, the face may be protected by a 'wear plate', which consists of a very thin wafer of quartz cemented to its surface. Good contact between the transducer and the workpiece is of paramount

importance, and some form of 'coupling' is required. The nature of the coupling is determined by the application. On smooth surfaces, water with a suitable wetting agent may be used. Other liquids include glycerine, oil, and, when the vessels or pipes are hot, a silicone of suitable specification. For very corroded surfaces, e.g., plates of ships that have seen service, prior surface preparation may be required. For curved surfaces, a crystal with a curved surface must be used. Experience has shown that with a flat crystal the radius of curvature of a convex workpiece must not be less than 3 in., and not less than 24 in. for a concave surface, so crystals have to be specially ground to deal with pipes having an outside diameter of less than 3–4 in. For small scale work barium titanate transducers are particularly suitable.

In the foregoing, the emphasis has been on metals, but certain of these do present difficulties. A case in point is cast-iron where porosity in the metal results in 'scatter'. The difficulty is increased by large grain size and variations in quality. Another difficult metal is lead, since vibrations are readily absorbed by it. To obtain accurate results, a high energy level is required for which a mains-operated meter must be used.

Among non-metallic substances that can be measured are glass and certain of the denser plastics. In all these measurements, it is essential that the velocity of sound in the material under test be accurately known. The necessary information can readily be ascertained in most instances – in extreme cases the instrument readings can be checked against test pieces of identical composition and of known thickness.

A point of particular practical importance is that, when measuring corroded and scaled surfaces, there is a tendency to get low readings because the signal is always reflected from the nearest face: from the safety point of view this is an advantage. On the other hand, if the surface is painted or otherwise protected, the thickness of the protective coat will not be measured.

Flaw Detection

As the interface between substances with different acoustical characteristics acts as a reflecting-surface for vibrations, ultrasonics may be used to detect flaws in materials; a crack, or change in composition of a material, forms a discontinuity in the path of the vibrations. While this method of non-destructive testing is not yet accepted by insurance companies, it may be used, for example, for the rapid testing of welds on high-pressure pipework and vessels. The method, in contrast to the conventional method of radiography, has the advantages of speed, portability, and absence of danger.

The pulse system is generally used in this application. Either two transducers are used, one as a transmitter and the other as a receiver, or one transducer which, by electronic switching, operates both as the transmitter and the receiver. The transmitted and received pulses deflect a synchronized spot on a cathode ray tube, the horizontal distance between the two peaks of deflection being proportional to the thickness of the material. A peak between these two points indicates a flaw and its position on the scale is a measure of the distance of the flaw from the surface of the material.

Other variations of this method entail the use of crystals placed on opposite sides of the material. Change in velocity of the vibrations, resulting from variations in the homogeneity of the material, caused by cracks, holes, and so on, is detected by the instrument. Altering the relative position of the transducers enables the direction of the flaw to be deduced. This method can be applied to various engineering materials, e.g., concrete.

Prevention of Scale

Mechanical vibrations may be used in boiler plant to prevent the formation of scale; the method used is to introduce ultrasonic waves into the feed water. By keeping the water and boiler in a state of constant agitation, deposition of scale is prevented and existing scale, if not too thick, will be loosened.

For this application a relatively low ultrasonic frequency is adequate, and an electro-mechanical (magnetostrictive) generator is quite suitable.

The circuit for energizing the transducer may consist of a condenser charged from a D.C. supply. At regular intervals the

condenser is disconnected from the supply and connected to the coil of the generator by means of a mercury switch. This sets up an oscillation in the circuit, the desired frequency of which (of the order of 30,000 cycles per second) is determined by the characteristics of the condenser and coil.

To avoid fatigue in the shell of the boiler, rivets, and so on, the wave amplitude is limited, i.e., a low energy level is used. One generator can feed up to say four oscillators – the number of oscillators required per boiler depends on its size. Energy consumption is in the region of 20 watts.

Soldering

An examination of the basic principles of soldering reveals that in this process also, ultrasonics can play a useful role. The essential requirement in a soldering process is to obtain a flow of solder over the two surfaces to be joined, the solder forming the bond between the two. The 'wetting' of the surfaces of the metal depends on their finish – they must be absolutely clean and free from oxide. To remove oxide and inhibit further oxidation, a suitable flux may be used, and the substance selected must be able to function at the temperatures required for soldering, since oxides continue to be formed during the process.

Many metals are easily tinned and soldered since their oxides can readily be removed by various compounds. Some metals, however, notably aluminium, present difficulty in this respect because their oxides are chemically inactive or tenaciously held. Some other method must therefore be adopted for oxide removal, and the principle of all these is mechanical abrasion under conditions which prevent the formation of fresh oxide. A method frequently used is to clean the surface of the metal with a wire brush while it is beneath a layer of molten solder. But such methods have obvious disadvantages, especially when dealing with thin delicate articles. The degree of bonding is also uncertain. On the other hand, ultrasonic vibrations enable the cleaning process to be performed with a high degree of efficiency throughout the soldering process.

When ultrasonic vibrations are applied to a liquid, they set up

rapidly fluctuating pressures which result in the formation of minute bubbles – a phenomenon known as cavitation. The instantaneous pressure set up round the bubbles is very large – in excess of the fatigue strength of a metal in contact with the bubbles. Consequently, minute pieces of metal are torn from the surface and flung into the liquid. This is equivalent to a cleaning action, and if applied to a metal surface capable of being tinned, a wetting action takes place.

For soldering, an electrically heated bit is secured to a magnetostrictive vibration generator, and the soldering is carried out in the conventional manner. (Inset 17.) A similar principle may be adopted for soldering baths, in which case the ultrasonic transducer is immersed in the bath.

Cleaning

The phenomenon of cavitation produced by ultrasonic vibration can also be used for the straightforward cleaning of materials and for the removal of dust and dirt from holes, slots, and similar inaccessible places. This method of cleaning is particularly useful for dealing with small and delicate articles, such as watch springs, hypodermic needles, and electric contacts. A recent application is the cleaning of articles contaminated with radioactive material. (Inset 18.)

The procedure generally adopted is to place the parts to be cleaned in a suitable cleansing solution, ultrasonic vibrations being applied by a transducer immersed in the solution. The surface of the transducer is vibrated at a sufficiently high intensity to produce the necessary cavitation on the surfaces of the parts to be cleaned. The resultant erosive action is equivalent to scrubbing, but it is much more efficient, since it penetrates even in the smallest crevices of the parts being treated, yet does not damage them. Moreover the time required is very much reduced.

Transducers are available for working temperatures up to about 170° F. Work at higher temperatures may be carried out by transmitting the ultrasonic vibration through a water bath with cooling coils. A magnetostrictive vibration generator is used, and self-contained units are available with the transducer

fitted in a small cleaning tank. Ultrasonic cleaning is possible with a wide range of frequencies, but those of the order of 40,000 cycles per second seem to be most satisfactory. This frequency provides efficient cleaning at low power levels with the absence of sonic and cavitation noises. Small changes in frequency have no significant effect on the cleaning operation.

Drilling

Another application of ultrasonics is to drilling and generally shaping very hard materials, such as glass, ceramics, etc. The tool of the required shape, which is vibrated at high frequency by a magnetostrictive generator, is applied to the material to be worked, and is 'lubricated' by an abrasive mixed with oil or any other suitable carrier. In this way, a hole is worn in the material, the abrasive particles embedding themselves in the tool which is not, therefore, subject to considerable wear. The tool itself is made of a tough but not brittle metal.

Conclusion

This article by no means exhausts all the possible applications to which mechanical vibrations may be put, and continuous development is taking place in this comparatively new field of applied physics. For example, the quality of mixing is improved by shaking. Consequently, very thorough mixing and degasifying of liquids can be obtained by high frequency vibration. At the present time, such equipment is only available for small mixes, but the liquids to which it is applicable include metal alloys. There are quite important possibilities in this development, since the strength of a material depends, among other things, on the homogeneity of its grain structure.

RESEARCH REPORT

A. W. HASLETT

SCREW-LIKE PARTICLES

PHYSICISTS at Columbia University have discovered a new property of nuclear particles: the association in some of them of a direction of rotation with a direction of movement, a property analogous to that of a screw. The discovery, confirmed within a few weeks at Chicago University, was the result of experiments suggested by two theoretical physicists, Professor T. D. Lee of Columbia University and Professor C. N. Yang of the Institute for Advanced Study, Princeton. Their suggestion arose out of a difficulty that had been encountered in explaining the different modes of breakdown shown by what appears to be a single kind of unstable meson particle – the τ-meson,* of 963 electron-masses. The evidence for the identity of these particles has become progressively stronger, as the values obtained for their masses and lifetimes have come more closely together. But to explain their different modes of breakdown, if they were really identical, was difficult or impossible without departing from ideas that had long been accepted in theoretical physics. Hence the suggested experiments and their confirmation – at first by two difficult experiments, but later more simply.

The first experiment was on the radioactive breakdown of oriented nuclei such that electrons are emitted when they break down. The nucleus that was suggested for this experiment, and used in it, was that of cobalt-60, and the experiment was done by Professor C. S. Wu of Columbia University, in collaboration with R. P. Hudson and others of the National Bureau of Standards. Quoting from a statement issued by the university: 'The

*The τ-meson was the first of these supposedly different particles to be described; they are referred to, as a whole, as K-mesons.

radioactive nucleus cobalt-60 was cooled to a temperature of 0.01° above absolute zero. At this temperature all thermal motions are reduced to extremely small values. The application of a magnetic field will cause most of the cobalt nuclei, which are known to be spinning, to align themselves like small magnets, parallel to the applied magnetic field. The radioactive nuclei disintegrate, giving off electrons. The crucial point is the comparison of the number of electrons emitted along the direction of spin to the number going in the opposite direction. The very fact that these numbers are different indicates the favouring of a direction associated with the spin, that is, a "handedness" in the sense of a screw.' The technical aspects, it was added, were 'quite difficult': 'Scintillation counters had to be installed within the complex vacuum and cooling system and extreme care had to be taken to eliminate spurious effects.' Outside the United States, the very low temperature needed for this experiment could have been produced only in Britain and Holland. Work in at least one other laboratory is known to be proceeding, but no results from it are yet available.

The second experiment was on the breakdown of μ-mesons into electrons. It involved the use of a high-energy accelerator, and was done by R. L. Garwin, Professor L. M. Lederman, and M. Weinrich of Columbia University. Again quoting from the university statement: 'It was discovered that when the familiar π-meson (well known since 1947, and now known to be principally responsible for the force that holds nuclei together) disintegrates into a μ-meson and a neutrino, the μ-meson always spins in the direction of its motion. Here again, the μ-meson advances as if it were a screw. . . . The alignment of μ-meson spins was detected by counting the end products of the radioactive μ-meson decay, again electrons, which were found to favour one direction of spin of the parent μ-meson over the other, in their direction of emission.'

In this experiment, as first done at Columbia, the directions of the emitted electrons were determined by counters. But this is not the only way of doing the experiment. It must have occurred at once to workers in many laboratories that the same result could be obtained by measuring up the angles between the tracks formed by μ-mesons and electrons in photographic emulsions already exposed in experiments with high-energy accelerators. This will probably have been tried; but there is a difficulty to be overcome, namely that the directional effect shown by the electrons may be masked by stray magnetic fields near the accelerator. The confirmation at Chicago, mentioned at the beginning of this note, was obtained by the emulsion method, but from a special experiment, done by V. L. Telegdi and J. I. Friedman. They put a triple-layer iron shield round the emulsion to reduce magnetic interference. Again the electron tracks showed a preference for one direction rather than the other.

The screw-like behaviour of particles which has been brought to light by these experiments is related directly to the analytical concept of 'parity'. This distinguishes between the ability or not to undergo reflection in a mirror – or in co-ordinate axes – without being changed in the sense that a mirror image is changed from the original, so that a distinction between left- and right-handedness is introduced.

Important as are the experiments described to theoretical physics, they refer only to interactions between particles in which the time-scale is long by nuclear standards, however short in terms of laboratory counters and circuits. Broadly, there is a division between interactions that take place on a time-scale of 10^{-10} seconds or longer, and are described as weak interactions, and those that take place 100,000 times more quickly, or still more so, and are described as strong interactions. Interactions of the first group include those of free meson particles and the type of radioactive breakdown in which electrons are emitted. Those of the second group include changes in the energy states of atomic nuclei. The distinction between these two groups may be one of kind or degree, so far as present knowledge takes us. It cannot be assumed, therefore, that the results of experiments done on one group of interactions have any direct relevance for the other.

With that caution, it appears that two distinct but related questions are involved. Of these, the one that, potentially, is

more practical is concerned with the representation of interactions between particles in wave-mechanical terms. The part of theoretical physics that tries to do so has been signally unsuccessful up to the present time in producing any usable results.* Evidently, if the particles show screw-like behaviour, so also must the 'operators' that represent their interactions. The point at issue is about the nature of the restrictions that are taken to be imposed. Restrictions in some form are the essence of any form of quantum theory. They apply to quantities which, in a closed system, remain unchanged in total amount and therefore are said to be conserved. Examples relevant to the properties of elementary particles are energy – which in this context means the mass of the particles when at rest – and the quantity of angular momentum or spin that the particle possesses. The latter, it may be noted, is sufficiently specified by one or other of two numbers ($+\frac{1}{2}$ and $-\frac{1}{2}$), and so carries less than the ordinary meaning of spin. Parity, in common with energy and spin, has been taken to be conserved in nuclear interactions; and, for the builder of theories, parity has been a yes-or-no property. This restriction is now removed, and there must be a new one to take its place. A mistaken foundation has thus been removed, and the way cleared for the finding of a new one that will serve better.

The second question that is raised relates rather to the philosophy of science. The interest, here, is that the laws of physics seem no longer to be unchanged by the substitution of left- for right-handedness and vice versa. This, at least, is the majority interpretation; whether it is logically necessary that, for example, left- and right-handed particles should exist in equal numbers may be another question. In any case Lee and Yang have pointed to a way out: that 'the overall symmetry of the universe may still be preserved by assuming that, if our galaxy is essentially right-handed, some distant galaxy may in turn be left-handed'. Neither the practical implications for the builders of theories nor the philosophical implications can yet be

*Some of the difficulties were discussed by G. Feldman in his articles in *Science News 35*.

assessed. But the builders of theories feel hopeful; they have been given a fresh start.

Lee, T. D. and Yang, C. N.: *Physical Review*, 1956 **104** 254.
Columbia University statement, January 15, 1957, published in full in *The New York Times* (New York and international editions, January 16, 1957), and in substance in *Science*, 1957 **125** 185.

HARDER THAN DIAMOND

The transformation of boron nitride, which in its usual state is a soft-feeling white powder, into a form that is both harder than diamond and more resistant to heat has been announced lately by the General Electric Company in New York. This seems to be the first case in which a material of new physical state, produced by ultra-high pressures, has potentially useful properties. Such methods were first used by Professor P. W. Bridgman of Harvard University. He worked at pressures up to 100,000 times that of the atmosphere; but he found only a few examples of a change of physical state that persisted after removal of the high pressure used to produce it. One of them was a 'black' form of phosphorus which he made from ordinary yellow phosphorus by exposing it either to pressures above 12,000 atmospheres at temperatures of more than 200° C, or to pressures of more than 35,000 atmospheres at room temperature. It was chemically inert, nearly half as dense again as yellow phosphorus, and electrically conducting. But there has been no suggestion that it is of use, or likely to be so.

Encouraged by Bridgman's researches, a team of physicists at the General Electrical Company's laboratories at Schenectady, New York, devoted four years of preparation to an attempt to produce synthetic diamonds. New forms of pressure chamber were devised, and advances made in the combination of high temperature and pressure. In one form of chamber, temperatures above 2,760° C were maintained at pressures of more than 100,000 atmospheres. Under these conditions, clusters of very small diamonds were produced in a few minutes. Larger, though still small, diamonds were obtained at a pressure of 53,000 atmospheres, maintained during some hours at an unstated tem-

perature. These results were announced in 1955. It was stated at the same time that experiments on the production of new materials were being undertaken.

An approach to the experiments on boron nitride may have been suggested by the periodic table of the elements. An atom of boron has one electron less in its outer shell than an atom of carbon, and an atom of nitrogen has one more. Hence there would be a general probability that a network of boron and nitrogen atoms might be formed under high pressure, in which the distribution of electrons would be similar to that in the all-carbon network of diamond. Possible further indications are given in the press statement issued in New York. In this, boron nitride was described in its normal state as 'white graphite', and the form of its crystals as 'hexagonal'. The structure of the low-pressure form of carbon, graphite, is a series of parallel sheets, each formed of a hexagonal network of carbon atoms. Hence the finding of a similar structure in the low-pressure form of boron nitride would have been a further encouragement to hope that its high-pressure form would resemble diamond.

The experiments were done by R. Wentorf, who had worked earlier on synthetic diamond. Under a pressure described as 'more than a million pounds per square inch' (i.e., more than 68,000 atmospheres) and at temperatures 'over 3,000° F' (more than 1,650° C), he produced crystals the size of grains of sand, which had a cubic form, 'strikingly similar to that of diamond', and which were hard enough to scratch diamond. Finally, whereas diamond begins to burn in air at a temperature of about 850° C, the high-pressure form of boron nitride withstands 1,900° C. The new material has been named 'Borazon'. From the figures of pressure and temperature quoted, it should be easier to make than synthetic diamond, and should have advantages in industrial use. Whether it will, in fact, be manufactured will presumably depend on the value of these advantages.

STRONTIUM-90 AND NUCLEAR EXPLOSIONS

The accumulation in human bone of strontium-90 formed in nuclear test explosions has appeared of exceptional interest

since the publication last year of the report of the Medical Research Council's committee on *The Hazards to Man of Nuclear and Allied Radiations*. One of the chief conclusions from this report was that the ratio of strontium-90 to calcium in bone might be expected to reach a level – which, on the standards proposed, would demand 'immediate consideration' – at a lower frequency of thermonuclear explosions than, on any reasonable assumptions, would be likely to produce genetic effects such as would be statistically detectable, even in the long run. This seemed to be a comforting conclusion, since the index of safety thus suggested was one that could be readily applied.

Further information on the accumulation of strontium-90 in bone has been given lately in a report by Dr W. F. Libby, one of the United States Atomic Energy Commissioners, to the National Academy of Sciences. This is of interest both for the conclusions to which this report leads, and in relation to the standards which he himself uses in discussion. In order to view this new report in perspective, it is necessary to return to the Medical Council's recommendations. These were based on an appendix to the British report, the authors of which were Professor W. V. Mayneord, of the Royal Marsden Hospital, London, and Professor J. S. Mitchell, holder of the Chair of Radiotherapeutics at Cambridge. Their argument began with the relatively slow development – often extending over 20 years later – of the changes in bone that are produced by radioactive material incorporated in it; such changes leading sometimes to bone cancer and leukaemia. For this reason, experience with radium, which extends over a full generation, has been taken as the best guide – in the first instance by the International Commission on Radiological Protection, and later when preparing the British report. The relative toxicities of radium and strontium-90 have accordingly been evaluated in experiments done on animals for the Commission, and on this basis a figure for the maximum permissible accumulation of strontium-90 in bone was arrived at; and this figure, like others, was stated specifically to be applicable to occupational workers.

The International Commission also made a general recom-

mendation in 1954 that in the case of the prolonged exposure of large populations, the maximum permissible levels should be reduced by a factor of 10 below those for occupational workers. At the Geneva conference on atomic energy, W. Binks, the secretary of the Commission, stated further that 'large populations' should not, for this purpose, be taken to mean more than 10 per cent of the total population; or, alternatively, that the maximum permissible dose should be further reduced. Arguments used by Mayneord and Mitchell in their appendix to the British report were that occupational workers are subject to special supervision and medical care; that rapidly growing bones, as in young children, may be expected to be more vulnerable to the effects of radiation; and that statistically small effects are more important in large populations. Emphasizing the many uncertainties involved, they concluded that a dosage equal to one-hundredth of that permissible in occupational workers would be unlikely to produce detectable effects even in whole populations; and that if the level of strontium-90 in bone showed signs of rising much above this level, there would be cause for immediate consideration.

The first point which arises from Dr Libby's report is that, even in the absence of further thermonuclear tests explosions, the level of strontium-90 in bone which according to the British report should demand attention is likely to be approached, if not reached, in the early 1970s. Dr Libby's estimate is, in fact, that 'the body burden in the United States' would amount, by that time, to about four-tenths of the M.R.C.'s safe level, and might possibly reach this level; while, in other countries, the effect of environmental factors might be to increase the level above that in the United States by not more than twice or three times, although the evidence as a whole suggested that so big an excess was unlikely. In view of these forecasts, there is an obvious need for further and up-to-date information.

A second point is about the standards cited both in Dr Libby's report, and also in a more recent American publication, judging from published Press summaries and comments in the case of the letter. Dr Libby's report refers throughout to the International Commission's recommendations for occupational workers as the 'maximum permissible concentration', and makes no mention of recommendations for a reduced maximum when applied to large populations. Further, the American practice is to average the total quantities estimated for the complete human skeleton, whereas a young child, with a smaller skeleton, must clearly be allowed a smaller quantity of radioactive material. It will be useful, as Dr Libby suggests, if the United Nations Scientific Committee on Effects of Atomic Radiations will undertake a co-operative study of the problem. It is important in the interval that a scientific attitude should be taken to standards, while admitting that the quantitative data behind them are limited.

Binks, W.: *Peaceful Uses of Atomic Energy: Proceedings of the International Conference in Geneva, August 1955*, 1956 **13** 129.

Libby, W. F.: *Proceedings of the National Academy of Sciences*, 1956 **42** 945.

Mayneord, W. V. and Mitchell, J. S.: *The Hazards to Man of Nuclear and Allied Radiations*, Appendix N. H.M.S.O. (1956).

'Recommendations of the International Commission on Radiological Protection', *British Journal of Radiology, Supplement No. 6* (1955).

MAGNETIC OBSERVATORIES

There are three observatories in Great Britain at which magnetic recordings are made. The oldest is in the control of the Astronomer Royal, and, like the astronomical observatory, it was first established at Greenwich where magnetic records have been kept since 1844.

By the early 20s, however, the railway electrification in the neighbourhood of Greenwich made a move imperative, and the observatory was re-established at Abinger, near Dorking, in Surrey. But now successive railway electrifications in that neighbourhood have driven this section of the Royal Greenwich Observatory to Hartland in North Devon. There is no railway nearer than twelve miles to Hartland, which was thought to offer the best prospect in southern England at the present time. Building of the new station at Hartland has been proceeding

during the past year, and observations were begun in January 1957. It will, however, be some time before the station is fully operative, and a period of overlapped observation will be necessary.

It is difficult to realize that until the early years of the present century, even Kew was sufficiently remote to be suitable for a magnetic observatory. But the advent of electric traction there brought about a serious deterioration and, in 1908, a new observatory to continue the work of Kew was built at Eskdalemuir, 800 feet up in the sheep country of north-east Dumfriesshire. For the first two years of its existence, Eskdalemuir was controlled by the Director of the National Physical Laboratory, but in 1910 it became the responsibility of the Directory of the Meteorological Office. A second observatory under the same control was set up at Lerwick in the Shetland Islands in 1923. Both are still in operation and have survived the advance of civilization.

CORRESPONDENCE

LITERARY STATISTICS – A SEQUEL

The Editors' attention has been called to an interesting sequel to an earlier article published in 'Science News'.

DR C. B. WILLIAMS, then head of the Department of Entomology at Rothamsted Experimental Station, contributed to *Science News No. 24* an article describing techniques of settling questions of disputed authorship by statistical methods. About 1942 he had hit upon the idea of determining authorship by differences in frequency-distribution of longer and shorter words. Later he learned that the late Dr Udney Yule, of Cambridge, had developed a similar technique based on length of sentences.

The idea, however, was much older than Dr Williams realized, when he wrote his *Science News* article and he has now published in *Biometrika* an account of work done on the subject by an American, Thomas Corwin Mendenhall, from 1880 onward.

How Dr Williams came to learn of Mendenhall's work is a story with a moral for any reader of this publication who, lacking scientific qualifications himself, comes across something that may be of interest to the expert. Rushworth Fogg, a Glasgow journalist, read in Calvin Hoffman's book *The Man Who Was Shakespeare* a somewhat vague account of statistical evidence purporting to favour Hoffman's thesis that Marlowe wrote the plays attributed to Shakespeare. Having had some controversy with Mr Hoffman at a press conference, Mr Fogg wrote him and asked for the source of his quotation but received no reply. He then wrote and asked Ohio State University, because Hoffman had mentioned that Mendenhall was a professor in the college now bearing that name.

Miss Ruth M. Erlandson, Reference Librarian, not only informed him that Mendenhall's article 'A mechanical solution of a literary problem' appeared in the *Popular Science Monthly*

in December 1901, and added that the periodical was available in the Mitchell Library in Glasgow.

This proved to be correct. From the point of view of furthering the argument Fogg was disappointed to discover that Hoffman had been accurate in reporting Mendenhall's technique of graphing frequency-distribution of word length, and also his comment that: 'In the characteristic curve of his plays Marlowe agrees with Shakespeare about as well as Shakespeare agrees with himself.' Not having material for further argument, Mr Fogg thought Dr Williams would probably know all about Mendenhall and wrote asking for details. The existence of Mendenhall was news to Dr Williams, who secured a photostat of the *Popular Science Monthly* article, and has now applied modern statistical method to this early American work.

His article in *Biometrika* refers to Mendenhall's acknowledgement in an earlier paper, published in 1887, of a suggestion by Augustus de Morgan, Professor Mathematics at University College, London, that it might be possible to identify an author by the average length of words he uses. Despite the fact that 'the mathematics of the comparison of frequency distributions was very little understood at the time when he was writing', Mendenhall triumphantly passes Dr Williams's tests.

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The Editors regret that space does not permit the publication of further comments arising from Bernard Mayo's articles on Theoretical Entities (*Science News 32* and *39*) nor Mr Mayo's comments on those already published.

SOME BOOKS RECEIVED

WHAT IS SCIENCE? edited by James R. Newman (London, Victor Gollancz, 1956), pp. 493, 21s.

The title page of this volume gives the additional information that in it 'twelve eminent scientists and philosophers explain their various fields to the layman'. It seems, to the reviewer at least, that the main title is most unfortunate. Isn't there something off-putting about a question that smacks of some dry philosophical wrangle – especially when the question isn't answered by the book anyway?

Almost without exception the distinguished contributors to this excellent volume take science as given and go on to describe its present content and outlook. It, in fact, meets, in large measure, the need for an outline of modern scientific knowledge. The editor has been a member of the editorial board of *The Scientific American* since 1948 and he has brought together a team that would be hard to beat.

A dozen chapters range from mathematics and logic through evolution and genetics to a final speculation on science as foresight. The book, says the editor, is addressed to the general reader, but he won't need to be too general or he will find some parts rather heavy going. But no matter. Science is now so vast that we can make contact with it at many points. It is rapidly and profoundly changing our conception of the Cosmos; it is revealing the mainsprings of our often irrational behaviour. Unfortunately, as Mr Newman points out, science is not wisdom; it may save us from the pox but not from our own folly. But, is it too much to hope that it may help – even at the eleventh hour?

THE GALATHEA DEEP SEA EXPEDITION 1950–1952 described by Members of the Expedition (London, George Allen and Unwin, 1956), pp. 296, 40s.

The sea covers three-quarters of the globe; its depths exceed by a wide margin the height of the greatest mountains; its *average* depth is greater than any hill in Britain; and its floor is furrowed by trenches of breathtaking magnitude. Throughout the whole of the thousand million cubic miles of water that make up the seven seas life is found. As Ragnar Spärck says in his introduction: 'At first sight the idea

of studying oceanic fauna by lowering a few tiny nets and bags may seem absurd'; and it would be true if it were not that in the ocean we do not get the point-to-point variation in ecological balance that we get on land – in the sea there is great uniformity throughout very large volumes, particularly at the greater depths.

Till the middle of the 19th century marine research was confined to shallow water – down to perhaps 40–50 fathoms – and it was widely held that life did not exist at greater depths. Then came the classic cruise of the *Challenger* (1872–6) and our knowledge of living things was pushed down to 1,000 fathoms.

Since then three great deep-sea expeditions have explored the most inaccessible part of the planet: the Danish World Exhibition of 1928–30 in the *Dana*, the Swedish *Albatross* Expedition of 1947–8, and now the Danish *Galathea* Deep Sea Expedition of 1950–2. So Denmark has made an unequalled contribution to our knowledge of the deep ocean, particularly through this last expedition equipped with all the aids of modern oceanographic research.

The aim of the expedition was to explore the deepest deep ocean and to discover what conditions of life existed there – and the depths now accessible to modern science go down to nearly 6,000 fathoms. As a result of *Galathea*'s investigations we now know that there is life at the bottom of the great trenches of the Pacific where it is absolutely dark, where the temperature is only just above freezing point, and the pressure is about 1,000 atmospheres.

Those who want to find out how these discoveries were made will find the preparation and history of the expedition most attractively presented by the members of the expedition.

CONTEMPORARY PHYSICS by C. F. von Weizsächer and J. Juilfs (London, Hutchinson's Scientific and Technical Publications, 1957), pp. 150, 18s.

Anyone looking for a guide to physics that is neither too long nor too difficult will find Weizsäcker and Juilfs's volume useful indeed. The content and achievement of physics, both classical and contemporary, are painted in broad outline without recourse to mathematics, and the level is suitable for the educated general reader who is prepared to do a bit of hard thinking, or the sixth-former who is in need of a glimpse of the wood after too much looking at the trees.

The authors do not see the volume as an end in itself, however. They know the limitations of predigested pabulum, and they cherish the hope that their efforts may awaken an honest hunger for more nourishing intellectual fare.

While any volume with such a title as this must of necessity give special prominence to the developments of the last 30 years, the classical concepts from which the contemporary have sprung are not neglected, and even those who were bred to classical physics, so to speak, will find a perusal of this succinct volume worth while, since, to quote the authors: 'today, even classical physics is viewed in a light different from that in which its originators would have thought possible'.

The book contains a sizeable list of recommendations for further reading, but the unwary lay reader is in for some shocks if he isn't careful in his follow-up since some of the great classical texts are included – texts that were difficult even to practitioners in the field when they first appeared. There is a short index, from which the omission of the name of 'J. J.' will seem strange to English eyes.

SCIENCE AND HUMAN LIFE by J. A. V. Butler (London, Pergamon Press, 1957), pp. 163, 15s.

Professor Butler's wide experience is reflected in this stimulating book. His early thermodynamic studies might have led one to expect that he would remain in the 'physical half' of science, but a Rockefeller Foundation Fellowship, which he held at the Rockefeller Institute of Medical Research at Princeton, introduced him to the biological side, and he is now professor of physical chemistry at the Institute of Cancer Research in the University of London. He has already tried his hand at interpreting modern biochemical discoveries to the non-processional reader, and now in *Science and Human Life* he discusses the significance of scientific discovery in relation to prevalent ideas of human nature.

The first half of his book is concerned with science and the individual, the second with science and human society. Different chapters are headed: 'Life as a Chemical Phenomenon', 'Pictures in the Mind', 'Primitive Beliefs', 'Science and Ethics', and 'Imagination in Human Life', each of which is a model of lucid exposition unburdened by masses of technical detail.

In clear and simple language that anyone can understand, Professor Butler discusses the evolution of scientific views of life in general, and human life in particular, and he analyses the meaning and limitations of the various views. What is particularly valuable is his

attempt to see the scientific view in the wider context of the age-old and universal attempts of all human societies to discover for themselves some meaning in their experiences and to formulate a working theory of life.

There is a useful list of books for further reading, which reflects the catholicity of Dr Butler's own studies, but there is no index. Between costing and publication the publishers have knocked 3s 6d off the price, an estimable idea well worth passing on.

THE ORGANISATION OF SCIENCE IN ENGLAND by D. S. L. Cardwell (London, Heinemann, 1957), pp. 204, 18s.

It took the first world war to reveal to the English that, up to that time, *laissez faire* had determined the development of science in England: it took the second to make quite clear to the public conscience that, for better for worse, we are now living in the age of science if not in a scientific age.

In a comparatively short space of time, a leisured amateur activity has changed into one of recognized professionalism. In an attempt to discover how these changes have come about, Dr Cardwell examines the social origin of applied science, and describes scientific education and technological progress in Europe from the 18th century. The book contains a mass of useful data, though the analysis of them is rather slight. Each chapter carries a long list of references and there is in addition a long list of nearly seventy works, quite a number of which are the more valuable for their being in the main unfamiliar.

GALACTIC NEBULAE AND INTERSTELLAR MATTER by Jean Dufay (London, Hutchinson's Scientific and Technical Publications, 1957), pp. 352, 60s.

Astrophysical research of the last quarter of a century has made nonsense of the phrase 'empty space'. Space is anything but empty, and we now know that possibly half the matter in our Galaxy exists in the form of interstellar gas or dust. From this diffuse matter it now seems probable that stars are formed continuously by a process of condensation, first into grains, and then into larger masses.

The speed of advance in this field of astrophysics during the past 25 years has been little short of astonishing and the investigation of interstellar matter is now of great interest. This book attempts to give an account of the many different observational techniques and

theoretical methods involved in such studies, and to trace the history of how the evidence for the existence of interstellar matter has been built up. Not only have we had to abandon the idea of the emptiness of space, but developments in radio-astronomy, and the discovery of the decimetric radiation emitted by the hydrogen atoms of space, have brought about profound changes in our conception of the structure of the Universe itself.

The book will appeal to two classes of reader: astronomers, and those who find modern astronomy 'fascinating for its own sake'. Useful bibliographies and indexes are included, and there are many excellent stellar photographs from the Observatories of Haute-Provence and Lyons, of which the author is Director.

A TEXTBOOK OF PLANT VIRUS DISEASES by Kenneth M. Smith (London, Churchill, 1957), pp. 652, 65s.

This is a volume to which the oft-misused adjective monumental may be applied with justice. It is a second edition brought up to date after twenty years. In these two decades our knowledge of virus diseases has expanded enormously and to make this information available the volume includes a section on every known disease affecting plants, fruit, and shade trees anywhere in the world.

If the properties of the virus are known they are given, as well as fully cross-referenced descriptions of the diseases they cause. The material is arranged alphabetically and there is a bibliography containing about a thousand items.

While the book is clearly written principally for the plant virologist, it should also be of immense value to the commercial fruit and vegetable grower who approaches his work in a scientific manner. To the amateur looking at the hundred-odd plates it may appear that all the plants he grows are infested with viruses – and probably they are!

ABOUT OUR CONTRIBUTORS

F. BRAILSFORD holds the degree of Doctor of Philosophy (Engineering) of the University of London, and is a member of the Institution of Electrical Engineers, and of the American Institute of Electrical Engineers. He was born in 1903 in Derbyshire, and served an appenticeship at H.M. Dockyard, Pembroke, followed by three years as a student in the University College of Swansea. After graduation he joined the Metropolitan-Vickers Electrical Company at Manchester and eventually took charge of the Electrical and Magnetic Section of the Research Department of that Company. Since 1951 he has been Professor of Electrical Engineering at University College, London.

D. J. BRUCE is a lecturer in the Department of Psychology at the University of Reading. His current interests include the effects of context on auditory and visual communication, and the application of cinematography to psychological work. He is married and has two children.

D. B. CARLISLE was educated at Manchester Grammar School and Magdalen College, Oxford, where he read zoology during the last years of the war. A period of research on the pituitary body of mammals, during which time he was Senior Demy of Magdalen College, was followed by two years at the Naples Zoological Station as Oxford Naples Scholar, Beit Medical Research Fellow, and finally as temporary head of the department of physiology. While in Naples his interest turned to the endocrinology of Crustacea, which is now his main field of research. For the past five years he has been endocrinologist at the Plymouth Laboratory of the Marine Biological Association of the United Kingdom. He has been managing editor of the *Pubblicazioni della Stazione Zoologica di Napoli* and is now sub-editor of the *Journal of the Marine Biological Association.*

M. C. H. DODGSON was born in 1919 and educated at Sherborne School, Dorset. He received his medical training at St Thomas' Medical School, qualifying in 1943 and taking M.D. in pathology in 1950. He spent 1944–6 as a Captain in the R.A.M.C. On returning from the Army he went to the Pathology Department of St Thomas' Hospital Medical school, where he did some teaching and started research on the microscopical structure of the brain, work which he continued in Bristol. Is now engaged in hospital pathology in the North of England.

A. LINFORD graduated from the University of London with a degree of Bachelor of Science in Civil Engineering in the late 1920s. He is now 49. Became interested in measurement and automatic control in the days when this specialized branch of engineering was a novelty. For the past 25 years has been engaged in the application of measuring instruments and automatic controllers to widely varied industries. Has thus seen 'automation' develop from the inside. Has travelled extensively both in Great Britain and overseas.

R. S. NYHOLM was born in Australia, and first came to England in 1947 on a Research Fellowship. He spent three years back in Sydney as Associate Professor of Inorganic Chemistry before returning in 1955 to take up his present position as Professor of Chemistry at University College, London. Although keen on the whole of inorganic chemistry, he is specially interested in the shape of molecules and the study of complex compounds. In addition to other physical methods of investigation he has made a particular study of the application of magnetic measurements in inorganic chemistry. Married, with three young children, he devotes what little of his leisure the latter do not absorb to walking and cricket.

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