Chapter 3

OLD MATH AND RENEWED PHYSICS: KEYS TO ENGINEERING COLD FUSION

Cynthia Kolb Whitney
Galilean Electrodynamics

ABSTRACT

This chapter argues that Cold Fusion (CF) can become more understandable, and hence more amenable to Engineering, especially Control Engineering, if we will take account of appropriate redevelopments in Physics. These redevelopments become possible with fuller application of standard Mathematics. Of particular interest is the venerable protocol of formulating governing differential equations, identifying an appropriate family of solutions for the differential equations, and applying the boundary conditions that specify the particular problem at hand. This protocol leads to review of some results from Special Relativity Theory (SRT), which leads in turn to a better entrée into Quantum Mechanics (QM), and hence into Chemistry and Electro-chemistry, and finally into CF. We find the following practical recommendations re CF: 1) Let the heavy water rest quietly for a long time before attempting to elicit fusion events; 2) Consider both gravity and electrochemistry to anticipate how various ion species will distribute themselves in the water; 3) Study a range of possible arrangements of electrodes in the water to determine the optimum one.

INTRODUCTION

This Chapter builds upon a Chapter from my 2013 Nova book titled *Algebraic Chemistry: Applications and Origins.* [1] That Chapter was titled "Electro-Chemistry in Power Generation", and its subject was CF - the much-sought economically practical alternative to Hot Fusion (HF); i.e., thermonuclear fusion.

Fusion in general exploits the fact that Hydrogen atomic nuclei include not only Hydrogen H nuclei (one proton), but also Deuterium D nuclei (one proton and one neutron), and even Tritium T nuclei (one proton and two neutrons). In fusion, the neutrons are needed to create output Helium He nuclei.

HF exploits the phenomena of plasma state and magnetic confinement. In plasma state, nuclei are stripped 7naked of electrons. In magnetic confinement, the stripped nuclei can come together close enough that nuclear forces can overcome electrical repulsion and allow nuclear fusion, producing He nuclei, and releasing much energy.

CF exploits properties of water and electrochemistry. The heavy nuclei fuse at lower temperatures for reasons not yet completely understood. Viewed on the basis of old nuclei consumed, or new nuclei formed, CF can release equally as much energy as HF.

HF is easy to understand in terms of current-day Physics, but it is hard to produce, because a huge energy investment is required before any excess energy, i.e., energy beyond the input energy investment, can be harvested. By contrast, CF is difficult to understand in terms of current-day Physics, and without such understanding, its production remains mysterious and uncertain. As a result, we hardly even get to the issue of controlling it for commercial applications!

This Chapter argues that CF can become more understandable, and hence more controllable, if we will just review how theoretical developments in Physics happened historically, and then do the obviously indicated further developments.

Sometimes in reviewing history, the most intriguing events are the ones that inexplicably did *not* occur at the times when they clearly could have. In the case of electromagnetic signals as the foundation for much of twentieth century physics, several possible events did not actually happen:

1) In the early twentieth century, Albert Einstein was working at almost the same time on both the concept of the photon [2], which later led to

Quantum Mechanics (QM), and on the concept of signal, which was the basis for his development Special Relativity Thoery (SRT) [3]. But he did not connect the photon and the signal to each other.

2) In the middle of the twentieth century, Richard Feynman was developing Quantum Electrodynamics (QED) [4], which featured the concept of virtual photons as mediators of electromagnetic fields. But he did not then suggest reviewing SRT in light of the virtual-photon concept.

3) Throughout that same mid-century time frame, Claude Shannon was developing the founding concepts of present-day Information Theory (IT), conveniently collected in Brilloin's book [5]. IT was a powerful tool for wartime code breaking and code making, and for the post-war communication and computation industries. It really led to our current 'Age of Information'. But Shannon, did *not* suggest using IT to scrutinize the Signal concept at the foundation of SRT. Nor did anyone else at that time.

Why did none of these events happen? We do not know, and we cannot guess. It is, however, possible to remedy the situation now, and to do it in exactly the same way it could have been done so much earlier. It is all just Applied Mathematics.

In the late nineteenth century there was already a well-established protocol for dealing with problems involving differential equations. There are three steps:

1) Specify the differential equations;
2) Characterize the family of solutions that they allow;
3) Identify and use the boundary conditions that represent your particular problem.

Thus the recommended Applied Math approach consists of these tools: the governing differential equations (D), an appropriate family of solutions that satisfy the differential equations (S), and the boundary conditions (B) that specify the particular problem at hand. Together they make the triad DSB, and define the DSB Protocol for solving an Applied Math problem.

This Protocol is extremely well developed, widely used throughout Engineering Science and Applied Physics, and fully reliable. It should be applied to any problem that allows its application; it should *not* be skipped over arbitrarily.

Another aspect of recommended approach comes from even further back: the discipline instilled by the ancient Greek founder of Geometry, Euclid. [6] He began his development of Geometry with an absolutely minimal set of Axioms (noun of Greek origin) that are stated in the form "Let us Postulate… (verb of Latin origin; also used as a noun equivalent to 'Axiom' in modern texts). Axioms or Postulates are not to be multiplied beyond the minimum possible set; an extra Postulate can only be redundant, or worse: conflictive. And if conflictive, an extra Postulate will inevitably produce Paradoxes.

It seems unlikely that Nature actually *has* Paradoxes; it seems more likely that only misguided human beings can have Paradoxes. So let us always be parsimonious with Postulates.

SRT uses two numbered Postulates. The First Postulate comes from Galilean Relativity; it says that the Laws of Nature (the differential equations of Physics) are the same in all inertial frames of reference coordinates. The Second Postulate specifies that the speed of light is always the constant *c* with respect to an observer in any inertial reference frame.

This pair of Postulates *could* constitute *too many* Postulates. Note that the resulting SRT does indeed possess an enormous literature about Paradoxes! In response to the SRT Paradoxes, the present approach dispenses with SRT's Second Postulate, and relies instead on just the First Postulate and the venerable DSB Protocol. Let us see next what it reveals.

ON THE NATURE OF ELECTROMAGNETIC SIGNALS

The governing partial differential equations for an electromagnetic signal are Maxwell's four first-order coupled field equations. Jackson [7] gives Maxwell's equations in modern notation and Gaussian units as:

$$\nabla \cdot \mathbf{B} = 0 \ , \ \nabla \cdot \mathbf{D} = 4\pi\rho \ , \ \nabla \times \mathbf{E} + \frac{1}{c}\partial \mathbf{B}/\partial t = 0 \ , \ \nabla \times \mathbf{H} - \frac{1}{c}\partial \mathbf{D}/\partial t = \frac{4\pi}{c}\mathbf{J}$$

Here $\mathbf{B}$ is the magnetic field and $\mathbf{E}$ is the electric field. The constant $1/c = \sqrt{\varepsilon_0\mu_0}$, where ε_0 is the electric permittivity and μ_0 is the magnetic permeability. In free space, $\mathbf{D} = \varepsilon_0\mathbf{E}$, $\mathbf{H} = \mathbf{B}/\mu_0$, and the charge density ρ and the current density $\mathbf{J}$ are zero. Free space will be the case of interest henceforth in this paper.

It is possible to insert two of the four coupled field equations into the other two, and thereby create two un-coupled, second-order, wave equations:

$$\nabla^2 \mathbf{E} - \frac{1}{c^2} \partial^2 \mathbf{E}/\partial t^2 = 0 \ , \ \nabla^2 \mathbf{B} - \frac{1}{c^2} \partial^2 \mathbf{B}/\partial t^2 = 0 \ , \text{ where } 1/c^2 = \varepsilon_0 \mu_0$$

Waves were a well-known physical phenomenon even before Maxwell came along. So at the turn of the 20th century it was tempting to think about electromagnetic signals as similar to familiar solutions to these wave equations.

The most familiar type of solution that the un-coupled wave equations support is the infinite plane wave. That name implies a wave that extends to infinity in all three spatial directions, and exists forever in time. In that context, the parameter c is interpreted as the speed of light. Given the history, it was only natural to think of electromagnetic signals as just like infinite plane waves. That is what Einstein did in 1905. [3] It is also what Feynman did half a century later. [4]

Viewed from our vantage point here in the early 21st century, IT actually provides a clear disqualifier for the idea that c is the universal speed for all electromagnetic signals. The problem is that c comes out of the uncoupled wave equations and their infinite-plane-wave solutions. But an infinite plane wave cannot function as a signal, because it cannot convey any information whatsoever! It is no better than an everlasting constant. It is to communication what the bagpipe is to music: an everlasting drone – until some kind of structure is imposed.

Communication of information requires amplitude modulation, or frequency modulation, or on-off switching: some other kind of structure. An infinite plane wave does not have any such structure. The prototypical type of solution that provides the needed structure is the finite-energy electromagnetic pulse. The pulse is what we need to make a proper model for a signal.

The key thing about finite-energy pulses, as apposed to infinite plane waves, is this: under the action of the four coupled field equations, the pulses do not remain pulses; they evolve into spread-out wavelets. This process is irreversible. It does not happen with the two un-coupled wave equations because those equations are second order in time, and so are invariant under time reversal. So no irreversible evolution can be revealed with them.

Let us detail the evolution. Let the word 'pulse' be the short description for a field profile that is rounded on top and sloping down on the sides, and

fading gradually to zero; for example, a Gaussian function. The mechanism for the waveform development is that the spatial derivatives applied in Maxwell's equations change the original Gaussian pulse into successively longer wavelets consisting of successively higher order Hermite polynomials multiplying the original Gaussian. As a result, the energy in the wavelet gets more and more spread out along the propagation path.

For a specific illustration, let the field directions be x and y, which makes the propagation direction z. For the record, a Gaussian pulse looks like

$$G(z) = \frac{1}{\sqrt{2\pi}} \exp(-z^2/2) \quad , \quad \text{and a Hermite polynomial looks like}$$

$$H_n(z) = \frac{1}{G(z)} d^n G(z) / dz^n \quad , \quad \text{with} \quad n = 1,2,3,4,... \text{ being the order of the}$$

Hermite polynomial.

Figure 1 illustrates a Gaussian pulse and the derived Hermite polynomial that it generates after one complete cycle (*i.e.,* four steps) through Maxwell's first-order coupled field equations. The horizontal axis is the propagation direction, z, and the vertical axis is the field amplitude, $G(z)$ to start with, and $H_4(z) \times G(z)$ after four steps. Series 1 is the input pulse, and Series 2 is the developed wavelet.

Figure 1. Illustration of signal waveform development.
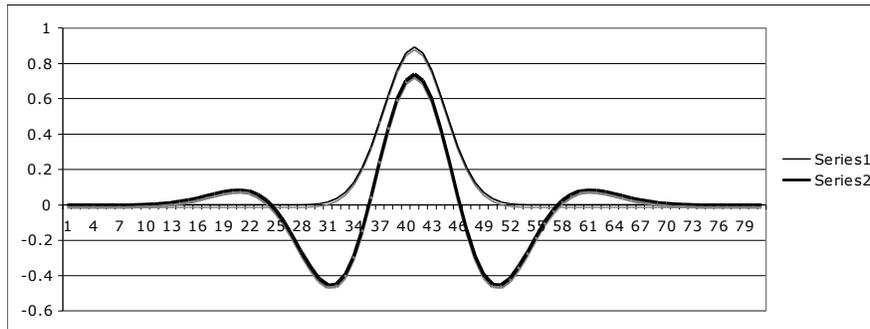
Now, to solve the stated propagation problem, one can pose a primary pair of pulses in **E** and **B** to guarantee travel, and, for more realism, add a second such pair, offset a quarter cycle in time and perpendicular in space, to model circular polarization, like a real physical photon exhibits. This amounts to four Gaussian functions. But that is not the end of it.

WHAT BOUNDARY CONDITIONS DEMAND

The third part of the Signal model is a pair of propagation boundary conditions: 1) no backflow of energy behind the source, and 2) no overflow of energy beyond the receiver. To guarantee these boundary conditions, we can demand $\mathbf{E} = \mathbf{0}$ at these boundaries. In Fig. 1, the boundaries correspond to the left and right borders. The field amplitude at the boundaries is so small that one cannot tell whether the boundary conditions have been imposed or not. Were the pulse to be near the beginning, or the end, of the path, then the effect of the boundary condition at the beginning, or at the end, of path would become more visible. Indeed, the propagation process near a boundary is really controlled by the boundary condition at that boundary. The fact that the receiver has this important role in determining the overall evolution of the signal waveform reminds one of a curious aspect of QM: the so-called 'observer effect'. Like QM, the present signal model has an observer effect; namely, the boundary condition at the receiver. But the signal model *also* has an equally important *source* effect; namely, the boundary condition at the source.

We can fulfill both of the required by boundary conditions by matching the signal leaving the source toward the receiver with 1) another fictitious signal going from the source in the opposite direction, with opposite $\mathbf{E}$, to make the total $\mathbf{E}$ field zero at the source, and 2) another fictitious signal approaching the receiver from the opposite direction, with opposite $\mathbf{E}$, so as to make the total $\mathbf{E}$ field zero at the receiver. One can even continue this boundary-fixing process, to clean up each tiny new departure from zero $\mathbf{E}$ at a boundary that each additional fictitious signal creates at the end of the path opposite to the end that it is meant to correct.

Carried to an infinite sum of corrections upon corrections, this is certainly a very complicated picture about an electromagnetic signal. How should we best characterize it in brief? We have all been conditioned to just give the number c for the speed, and presume that number to be relative to the 'observer'; i.e., the receiver of the signal. But clearly, that characterization is inadequate. When we speak of *the* speed of light, we speak way too casually. The little article word '*the*' implies way too much: that there is just *one* light-speed, the number c. But when a light pulse not only travels, but also evolves, there is not a single light-speed *number*. There is a light-speed *function*, depending on both position and time.

It is possible to think about this light-speed function in terms of local Poynting vector $\mathbf{P} = \mathbf{E} \times \mathbf{B}$ and local energy density $S = (E^2 + B^2)/2$. One can speak of a local light speed $c(z) = P(z)/S(z)$. One can speak of a path averaged light speed $<c> = \int P(z)\,dz \Big/ \int S(z)\,dz$. One can include the time variation in $c(z,t) = P(z,t)/S(z,t)$ and $<c>(t) = \int P(z,t)\,dx \Big/ \int S(z,t)\,dz$. One can recognize that $\int S(z,t)\,dz$ is total energy and should be independent of time. One can intuit that while most of the energy is near the source, $<c>(t)$ is c relative to the source, and while most of the energy is near the receiver, $<c>(t)$ is c relative to the receiver. But what about the great bulk of the scenario, during which the energy is *not* near either the source *or* the receiver?

We need a new and more appropriate statement to use for the DSB boundary conditions. The above statements can be characterized as 'one number, two references'. Instead, let us consider the reverse: 'one reference, two numbers'. First, look at the propagation scenario overall, and take note of an important symmetry property that it exhibits. There exists a temporal midpoint of the scenario, and at that moment, the waveform is symmetric about the spatial midpoint of the propagation path. Figure 1 corresponds to this mid-point situation, with the source on the left side of the Figure, and the receiver on the right side.

Consider that it cannot matter very much how the receiver *previously* moved before that temporal midpoint, and it cannot matter very much how the source will *subsequently* move after that temporal midpoint. The signal is roughly like a baton, handed off from source to receiver at that temporal midpoint. The length of that baton defines the distance traveled from the source to the receiver, and the direction of the baton defines the angular position of the source in relation to the receiver.

A good Rule of Thumb is the following: Whenever we cannot avoid speaking of light speed as a number c relative to something, let that something be the spatial midpoint of that imaginary baton. This statement is different from Einstein's Second Postulate [3] that c be always relative to the receiver, and different from Ritz's counter proposal [7] that c be always relative to the source, *and* different from the first option described above: 'one number, two references'. When we speak of c relative to the spatial midpoint of the propagation scenario, that characterization amounts to *not* c relative to

the source and *not* c relative to the receiver. It is the realization of the second option: 'one reference, two numbers'.

To be sure, the Rule of Thumb offered here is an approximation. But absent large accelerations, it is a very good approximation. So let us see next what new understanding it produces.

REALITY *VS*. INTERPRETATION

Consider a simple scenario with steady relative motion between an observer and a target. Consider a real-world problem: estimating the speed of a moving target relative to an observer. One way or another, there must exist *two* signals, the first one outgoing from the observer to the target, and the second one returning from the target to the observer. So for one signal, the observer is the source and the target is the receiver, and for the other signal, the target is the source and the observer is the receiver.

Each of the two signals has two phases: expansion from its source, and collapse to its receiver. Thus this scenario has altogether *four* parts. Call them Parts 1 through 4.

If the motion is retreat at speed V, then having speed c relative to the scenario midpoint means we have $c + V / 2$ relative to the source, and $c - V / 2$ relative to the receiver (presuming for now that $V < 2c$). If the motion is approach at speed V, we have the opposite case, $c - V / 2$ relative to the source (again presuming $V < 2c$), and $c + V / 2$ relative to the receiver.

To keep track, it helps to retire the vague concept of '*the* speed of light', and speak more specifically of 'the speed of the leading front of the developing wavelet' and 'the speed of the trailing end of the developing wavelet'. During Part 1, the outgoing expansion from the observer, the leading front moves at $2c$ relative to the source, or $2c - V$ relative to the target. During Part 2, the collapse to the target, the trailing end moves at $2c$ relative to the target, or $2c + V$ relative to the source. Then during Part 3, the return expansion, the leading front moves at $2c$ relative to the target, or $2c - V$ relative to the observer. Finally, during Part 4, the collapse to the observer, the trailing end moves at $2c$ relative to the observer, $2c + V$ relative to the target.

Suppose the initial separation between source and target is L_0, and the relative speed between source and target is V. For the record, if $V \geq 2c$ the linear-motion scenario stops right here: either the signal from the observer could never reach its target, or the return signal from the target could never

arrive back to the observer. So we continue the present analysis under the assumption that $V < 2c$.

1) After Part 1 (expansion from the observer) the separation is:

$$L_1 = L_0 + L_0 \times (V / 2c) + L_0 \times (V / 2c)^2 + ... = L_0 / (1 - V / 2c) .$$

2) After Part 2 (collapse to the target) the separation is:

$$L_2 = L_1 \times (1 + V / 2c) = L_0 \times (1 + V / 2c) / (1 - V / 2c) .$$

3) After Part 3 (re-expansion from the target) the separation is:

$$L_3 = L_2 + L_2 \times (V / 2c) + L_2 \times (V / 2c)^2 + ... = L_2 / (1 - V / 2c)$$
$$= L_0 \times (1 + V / 2c) / (1 - V / 2c)^2 .$$

4) After Part 4 (re-collapse to the observer) the separation is:

$$L_4 = L_3 \times (1 + V / 2c) = L_0 \times (1 + V / 2c)^2 / (1 - V / 2c)^2 .$$

Compare this case to the case where the scenario follows the Einstein Postulate. There would be only two parts: outbound travel at c relative to the target, followed by return travel at c relative to the observer. And the scenario could not play out if $V \geq c$. So we continue the analysis under the assumption that $V < c$.

1) After Part 1 (travel from the observer to the target) the separation would be:

$$L_1 = L_0 / (1 - V / c) .$$

2) After Part 2 (return travel from the target to the observer) the separation would be:

$$L_2 = L_1 \times (1 + V / c) = L_0 \times (1 + V / c) / (1 - V / c) .$$

Observe that the length expansion with the Einstein Postulate is less than that with the four steps.

For the sake of completeness, consider also the case where the scenario follows the Ritz Proposal. Again there would be only two parts, but they would be slightly different, because the outbound trip would be at c relative to the observer, and the return trip would be at c relative to the target.

1) After Part 1 (travel from the observer to the target) the separation would be:

$$L_1 = L_0 \times (1 + V/c) \ .$$

2) After Part 2 (return travel from the target to the observer) the separation would be:

$$L_2 = L_1 \big/ (1 - V/c) = L_0 \times (1 + V/c) \big/ (1 - V/c) \ .$$

So the length expansion with the Ritz Postulate is similarly less than that with the four steps that make the present model.

These discrepancies create a problem. When the actual function is $(1 + V/2c)^2 \big/ (1 - V/2c)^2$, while the believed model is $(1 + V/c) \big/ (1 - V/c)$, the only way to make the believed model match the actual function is to believe in a reduced value of speed, some $v < V$. To produce the required match between the actual function and the believed model, one needs $v = V \big/ (1 + V^2/4c^2)$:

$$\frac{1 + v/c}{1 - v/c} = \frac{1 + (V/c) \big/ (1 + V^2/4c^2)}{1 - (V/c) \big/ (1 + V^2/4c^2)} = \frac{1 + V/c + V^2/4c^2}{1 - V/c + V^2/4c^2} \equiv \frac{(1 + V/2c)^2}{(1 - V/2c)^2} \ .$$

Given the distortion $v = V \big/ (1 + V^2/4c^2)$, other distortions quickly follow. If the data being collected concerns the length of a measuring rod, then the rod appears to be shortened in the travel direction, or if the data concerns the image of a clock face, then the clock appears to run slow, and so on.

Observe that the function $\dfrac{1+v/c}{1-v/c}$ blows up for $v = c$, which is surprising. Further observe that the function $\dfrac{1+v/c}{1-v/c}$ changes sign for $v > c$, which makes no physical sense at all. Because of these facts, it has long been thought that Nature has a physical speed limit $v < c$. But the idea that Nature would have a speed limit is also generally recognized as a Big Mystery.

The present analysis suggests that little v is just a convention, an artifact accepted from a somewhat faulty mathematical model, and *not* a fact established from unbiased observation or uncontested theory. Situations like this are well known in Engineering Science: in controlling the behavior of a physical system, a mathematical model that strays too far from the physical reality will lead to strange, and sometimes catastrophic, consequences.

But what about big V? Observe that the function $\dfrac{(1+V/2c)^2}{(1-V/2c)^2}$ blows up for $V = 2c$. Does this fact mean that V has a limit $V < 2c$ corresponding to $v < c$? No, it does not! What V has is a limit on *observation*: speeds $V > 2c$ may indeed exist, but for linear motion, they simply cannot be observed using light signals.

In circular motion, however, the situation is different. Speed $V > 2c$ may be observable, and it may, in fact, have been observed. Kuligin et al., [9] tell a story suggesting this conclusion. Let a linear particle accelerator deliver a bolus of charged particles into a magnetic field, where charges follow a circular orbit.

The resulting situation sometimes *looks* as if the charge bolus breaks into fragments. But the observation *also* allows the alternative interpretation that the charge bolus remains intact, but travels *much faster* than SRT allows. So in a time interval for which we expect to see the charge bolus go by only once, we actually see it go by twice, or even more times!

Now for the sake of completeness, observe that the relationship $v = V\big/(1+V^2/4c^2)$ implies a quadratic equation for V as a function of v:

$$(v/4c^2)V^2 - V + v = 0 \ .$$

This equation has two solutions, $V = \dfrac{1 \pm \sqrt{1 - v^2 / c^2}}{v / 2c^2}$. For small v ,

$V = \dfrac{1 \pm (1 - v^2 / 2c^2)}{v / 2c^2}$, which evaluates to $V = \mp v$ and $V = \dfrac{2 - v^2 / 2c^2}{v / 2c} \to \dfrac{4c^2}{v}$.

That is to say, V evaluates to one familiarly small value, and one un-familiarly huge value!

POTENTIALS AND FIELDS

Let us now proceed to review SRT in a different and more important matter; namely, electromagnetic potentials and fields. These are functions of position $\mathbf{r}$ and time t where some receiver can sense them. The potentials are the scalar potential $\Phi(\mathbf{r},t)$ and the vector potential $\mathbf{A}(\mathbf{r},t)$. Through spatial and temporal derivatives, they give rise to the electric field $\mathbf{E}(\mathbf{r},t)$ and the magnetic field $\mathbf{B}(\mathbf{r},t)$:

$$\mathbf{E}(\mathbf{r},t) = -\nabla\Phi(\mathbf{r},t) - \frac{1}{c}\partial\mathbf{A}/\partial t \quad , \quad \mathbf{B}(\mathbf{r},t) = \nabla \times \mathbf{A}(\mathbf{r},t) .$$

The problem of characterizing the electromagnetic potentials and fields registered at a receiver predates Einstein, and the Assumption that all researchers made about it was so ingrained that it was not even recognized as being an Assumption, until Einstein called it out as his Second Postulate in [3]. So he is to be much credited for doing that.

The sources generally cited for the early (1898 to 1901) work on potentials and fields are A. Liénard [10] and E. Wiechert [11]. Although they worked at about the same time, they worked separately. And they got the same results – as did everyone else working on the problem around that time.

The Liénard-Wiechert results are given in every standard EM book, and are summarized again here, for review and subsequent further discussion. The standard scalar and vector potentials are:

$$\Phi(\mathbf{r},t) = e\left[1/\kappa R\right]_{\text{retarded}} \text{ and } \mathbf{A}(\mathbf{r},t) = e\left[\boldsymbol{\beta}/\kappa R\right]_{\text{retarded}} ,$$

where $\kappa = 1 - \mathbf{n} \cdot \boldsymbol{\beta}$, $\boldsymbol{\beta}$ is source velocity normalized by c, and $\mathbf{n} = \mathbf{R} / R$ (a unit vector), and $\mathbf{R} = \mathbf{r}_{\text{source}}(t - R / c) - \mathbf{r}_{\text{receiver}}(t)$ (an implicit definition for the terminology 'retarded').

The LW fields obtained from those potentials are then:

$$\mathbf{E}(\mathbf{r},t) = e \left\{ \frac{1}{\kappa^3 R^2} (\mathbf{n} - \boldsymbol{\beta})(1 - \beta^2) + \frac{1}{c\kappa^3 R} \mathbf{n} \times \left[(\mathbf{n} - \boldsymbol{\beta}) \times (d\boldsymbol{\beta} / dt) \right] \right\}_{\text{retarded}}$$

and $\mathbf{B}(\mathbf{r},t) = \mathbf{n}_{\text{retarded}} \times \mathbf{E}(\mathbf{r},t)$.

The $1/R^2$ fields are called Coulomb-Ampere fields. Observe the direction of the Coulomb electric field: $(\mathbf{n} - \boldsymbol{\beta})_{\text{retarded}}$. If $\boldsymbol{\beta}$ does not change much over the total field propagation time, then $(\mathbf{n} - \boldsymbol{\beta})_{\text{retarded}}$ is virtually indistinguishable from $\mathbf{n}_{\text{present}}$.

The $1/R$ fields are radiation fields, and together they make a Poynting vector (energy flow per unit area per unit time):

$$\mathbf{P} = \frac{c}{4\pi} \mathbf{E}_{\text{radiative}} \times \mathbf{B}_{\text{radiative}} = \frac{c}{4\pi} \mathbf{E}_{\text{radiative}} \times \left(\mathbf{n}_{\text{retarded}} \times \mathbf{E}_{\text{radiative}} \right) = \frac{c}{4\pi} \left(E_{\text{radiative}} \right)^2 \mathbf{n}_{\text{retarded}}$$

The Poynting vector lies along $\mathbf{n}_{\text{retarded}}$.

So the Coulomb field and the Poynting vector are arriving to the observer from distinctly different directions: approximately $\mathbf{n}_{\text{present}}$ *vs.* $\mathbf{n}_{\text{retarded}}$. This result does *not* look right: it looks like advance information being provided on one channel, but not the other, of the two information channels: the Coulomb signal and radiation signal.

Nevertheless, everyone working at that time adopted the same Assumption, and endorsed the same result. The Liénard-Wiechert Assumption, and later the Einstein Postulate, was: the speed of light is always c with respect to the receiver of the light.

Now consider the same problem using the new and more nuanced definition for the light speed reference: the 'baton' handed over from the source to the receiver at the temporal midpoint of the propagation scenario. Observe that the line that connects the source position to the receiver position at the temporal midpoint of the scenario defines both the distance and the

direction that the energy must travel in order to achieve the particular transmission of a signal from the source to the receiver that is being modeled. This specification implies that we need potentials and fields to be, not *retarded*, but ***half-retarded***. Now the potentials become:

$$\Phi(\mathbf{r},t) = e\left[1/\kappa R\right]_{\mathbf{half}-\text{retarded}} \text{ and } \mathbf{A}(\mathbf{r},t) = e\left[\boldsymbol{\beta}/\kappa R\right]_{\mathbf{half}-\text{retarded}} .$$

The fields become:

$$\mathbf{E}(\mathbf{r},t) = e\left\{\frac{(\mathbf{n}-\boldsymbol{\beta})(1-\beta^2)}{\kappa^3 R^2} + \frac{\mathbf{n}}{c\kappa^3 R}\times\left[(\mathbf{n}-\boldsymbol{\beta})\times d\boldsymbol{\beta}/dt\right]\right\}_{\mathbf{half}-\text{retarded}}$$

and $\mathbf{B}(\mathbf{r},t) = \mathbf{n}_{\mathbf{half}-\text{retarded}}\times\mathbf{E}(\mathbf{r},t)$.

Observe the direction of the Coulomb $\mathbf{E}$ field:

$$(\mathbf{n}-\boldsymbol{\beta})_{\mathbf{half}\text{-retarted}} \approx (\mathbf{n}_{\text{present}})_{\mathbf{half}\text{-retarted}} \overset{\triangle}{=} \mathbf{n}_{\mathbf{half}\text{-retarted}} .$$

The Poynting vector becomes:

$$\mathbf{P} = \frac{c}{4\pi}\mathbf{E}_{\text{radiative}}\times\mathbf{B}_{\text{radiative}}$$

$$= \frac{c}{4\pi}\mathbf{E}_{\text{radiative}}\times\left(\mathbf{n}_{\mathbf{half}-\text{retarded}}\times\mathbf{E}_{\text{radiative}}\right) = \frac{c}{4\pi}(E_{\text{radiative}})^2\mathbf{n}_{\mathbf{half}-\text{retarded}} .$$

This vector too lies along $\mathbf{n}_{\mathbf{half}-\text{retarded}}$. That is, the Coulomb field and the Poynting vector are now reconciled to the *same* direction, instead of conflicting with each other. Does this reconciliation really matter? Indeed it does. With this one little adjustment, we are far better prepared to understand the next problem at hand: the behavior of an atom.

MODELING AN ATOM

QM as we presently know it was founded, in part, to avoid a seeming problem with the Hydrogen atom viewed classically: the electron in its circular

orbit must accelerate, and so must radiate away its orbit energy, and so must collapse into the nucleus. That circumstance forces us to inject yet another Postulate; namely, that classical physics is not sufficient at the level of atoms.

But with the direction of the Coulomb field being $\mathbf{n}_{\mathbf{half-retarded}}$, there is a tiny tangential component of Coulomb force aligned with the orbit velocity. So there is a torque on the atom, and the torque pumps energy into the atom, and that process can work to balance the energy loss due to radiation. Here are some relevant details:

Let the masses of the electron and the proton be m_e and m_p . Note that $m_e \ll m_p$, but m_p is *not* infinite.

Let the orbit radii of the electron and the proton be r_e and r_p . Note that $r_p \ll r_e$, but r_p is *not* zero.

Let the charges on the electron and the proton be $-e$ and $+e$.

The magnitude of the nominally attractive force within the atom is $F = e^2 / (r_e + r_p)^2$.

Let the orbit frequency be Ω . The orbit speed of the electron is $v_e = r_e \Omega$ and that of the proton is $v_p = r_p \Omega$.

The magnitude of the tiny tangential force on the electron is $F_e = F v_p / 2c = F r_p \Omega / 2c$.

The magnitude of the tiny tangential force on the proton is $F_p = F v_e / 2c = F r_e \Omega / 2c$.

The magnitudes of the torques on the electron and on the proton are $T_e = r_e F_e$ and $T_p = r_p F_p$, both equal to $F r_e r_p \Omega / 2c$.

The total torque on the electron-proton system is $T_{total} = T_e + T_p = 2T = F r_e r_p \Omega^2 / c$.

The torque power delivered to the system is $P_{torque} = T\Omega = F r_e r_p \Omega^2 / c$.

The torque power delivered to the system is $P_{torque} = T\Omega = F r_e r_p \Omega^2 / c$.

The squared orbit frequency is determined from either $F = m_e r_e \Omega^2$ or $F = m_p r_p \Omega^2$.

The more convenient of the two options is $\Omega^2 = F / m_e r_e$. With that expression, the approximation $r_e \approx r_e + r_p$ yields:

$$P_{\text{torque}} = F r_e r_p \Omega^2 \big/ c = F^2 r_p \big/ m_e c = \left(e^4 r_p \big/ m_e c\right) \big/ \left(r_e + r_p\right)^4 \approx \left(e^4 \big/ m_p c\right) \big/ \left(r_e + r_p\right)^3$$

The P_{torque} is actually quite large, and so it changes the whole emphasis of worry concerning the Hydrogen atom. The question becomes, not why does the Hydrogen atom not radiate and collapse to death, but rather why does the Hydrogen atom not torque itself up and expand way beyond its known size?

The fact is: there exists much, much more radiation than was previously worried about, and it is enough to produce the proper balance between radiation and torquing. The extra radiation arises from finite signal propagation speed, which results in circular motion of the center of mass of the Hydrogen atom. This in turn produces Thomas rotation [12], and thereby scales up by a factor of 2 the overall rotation rate generating the radiation, which increases the radiation power by a factor of 2^4 .

But what should one say about that center-of-mass circular motion? In Newtonian physics, where signal propagation speed is infinite, there is no such thing. In Maxwell physics, the emphasis is on fields, and the responses of individual charges, but not as much on the responses of whole charge systems, such as atoms. So the issue doesn't come up there. In Einstein's relativity physics, the emphasis is often on the observers of events more than on the events themselves. System center-of-mass circulation seems not to come up, although Thomas rotation does.

Thomas rotation is generally believed to be a result of the properties of Lorentz transformations, and hence of SRT. That is the belief because one can think of Lorentz transformations, not only in the usual, passive sense, as conversion from an observer in one inertial coordinate frame to another observer in another inertial coordinate frame, but also in the active sense, as the application of a 'boost' in velocity; that is, as an acceleration, the result of a physical force. A series of non-co-linear boosts does indeed produce Thomas rotation.

But Thomas rotation does arise, not just from Lorentz transformations, but also from Galilean Transformations. That fact can be demonstrated in detail as follows:

For simplicity, let all motion be in the x, y plane. The scenario begins at time coordinate ct_0 with one of the particles, say the electron, at rest at spatial coordinates x_0, y_0. Let an attraction from another particle act in the x direction. Let an increment of velocity $\Delta\mathbf{V}_x = \Delta V$ be imposed, and let an increment of time Δt elapse. The coordinates of the electron then become:

$$t_1 = c(t_0 + \Delta t) \ , \ x_1 = x_0 + \Delta V \Delta t \ , \text{ and } y_1 = y_0 \ .$$

Now let an attraction from another particle act in the y direction. Let an increment of velocity $\Delta\mathbf{V}_y = \Delta V$ be imposed, and let another increment of time Δt elapse. The coordinates of the electron then become:

$$t_2 = c(t_0 + \Delta t + \Delta t), \ x_2 = x_0 + \Delta V(\Delta t + \Delta t) = x_0 + 2\Delta V \Delta t \ , \text{ and,}$$
$$y_2 = y_0 + \Delta V \Delta t \ .$$

Observe that, if the Galilean velocity boosts had been applied in the opposite order, then the ending coordinates of the electron would have been:

$$t_2 = c(t_0 + \Delta t + \Delta t) \ , \ x_2 = x_0 + \Delta V \Delta t \ , \text{ and } y_2 = y_0 + 2\Delta V \Delta t \ .$$

Observe that the squared incremental length changes have the same magnitude either way:

$$(2\Delta V \Delta t)^2 + (\Delta V \Delta t)^2 \equiv (\Delta V \Delta t)^2 + (2\Delta V \Delta t)^2 = 5(\Delta V \Delta t)^2$$

That fact means the two possible sequences of Galilean boost applications differ only by a rotation. That means each one individually contains a rotation equal to half that total angle difference. This is the Thomas rotation.

Without the Thomas rotation, the radiation power from the Hydrogen atom would be:

$$P_{\text{radiated}} = \frac{2e^2}{3c^3} a_{\text{e}}^2 = \left(2e^6 / m_{\text{e}}^2\right) \Big/ 3c^3 (r_{\text{e}} + r_{\text{p}})^4 \ .$$

With the Thomas rotation included, the radiation power from the Hydrogen atom is:

$$P_{\text{total radiated}} = 2^4 \frac{2e^2}{3c^3} a_e^2 = \left(2^5 e^6 / m_e^2\right)\Big/ 3c^3 (r_e + r_p)^4 \ .$$

The value of the separation $r_e + r_p$ for which $P_{\text{total radiated}} = P_{\text{torque}}$ is:

$$r_e + r_p = 32 m_p e^2 \Big/ 3 m_e^2 c^2 = 5.5 \times 10^{-9} \text{ cm} \ .$$

Recall that in the traditional approach to QM, $r_e + r_p = h^2 \big/ 4\pi^2 \mu e^2$, where μ is the reduced mass, defined by $\mu^{-1} = m_e^{-1} + m_p^{-1}$, and very nearly equal to m_e , and h is Planck's constant, 6.626176×10^{-34} Joule-sec, a fundamental constant given by Nature.

The present approach does not use Planck's constant as an input. Instead, it gives an estimate of Planck's constant as an output:

$$h = \sqrt{4\pi^2 \mu e^2 (r_e + r_p)} \approx \sqrt{4\pi^2 m_e e^2 \, 32 m_p e^2 \big/ 3 m_e^2 c^2}$$
$$\approx \frac{\pi e^2}{c} \sqrt{128 m_p / 3 m_e} \approx 6.77 \times 10^{-34} \text{Joule-sec.}$$

This estimate can be improved by taking due account of the fact that the sines of small angles are not exactly equal to the angles themselves, and the cosines of small angles are not exactly equal to unity. The sine corrections are third order in angle, while the cosine corrections are second order in angle, which is more significant. Including those corrections reduces the radiation power slightly, and so reduces the solution $r_e + r_p$ slightly, and so reduces the estimate of h slightly – a step in the right direction.

What is the broader significance of this analysis of the Hydrogen atom? It illustrates an approach that involves no founding Postulate, and requires no elaborate computation; it is just Algebra. The approach can help, not just with the Hydrogen atom, but also with *all* atoms. And it can help, not just with single atoms, but also with complex *systems* consisting of multiple atoms.

I see two major areas of study involving complex systems consisting of multiple atoms. One area is Spectroscopy. It has been suggested that *all* light generation may be similar to laser-light generation. [13] Laser light arises, *not* from *single* atoms, but rather from *multiple* atoms. The multiple atoms constitute a complex system, and the complex system emits *multiple* photons upon transition into other less complex systems. That idea can account for the complexity of observable spectra, more plausibly I believe, than can the traditional idea of single atoms possessing multiple possible excited states. This area of investigation is certainly interesting, but it is not obviously relevant to CF.

The other area of study involving complex systems consisting of multiple atoms is Chemistry. I believe a convenient approach to Chemistry is truly vital for understanding and engineering CF. The next Section discusses a new approach to Chemistry that offers a computational advantage over existing approaches.

THE BASICS OF ALGEBRAIC CHEMISTRY

The title of my book [1] Algebraic Chemistry (AC) reflects the fact that the technique has no integrals or other complicated math operations that would demand capabilities beyond those of a hand calculator. The worst operation is square root. So the AC approach looked promising for quick application to problems such as CF.

The fundamental idea behind AC is that all atoms share some similarities with Hydrogen atoms: 1) They have a nucleus that is similar to a proton, but scaled up to nuclear charge Z and nuclear mass M ; 2) They have a population of electrons that is not entirely unlike a single electron; i.e.*,* an interacting community that is somewhat coherent, and somewhat like one *big* electron orbiting the nucleus; 3) It is possible for the electron count to be different from the nuclear charge. This last possibility is what characterizes ions, and thereby creates all of Chemistry.

The initial observation from which the fundamental idea developed is this: neutrons are a complicating factor for Chemistry. They create a huge variety of slightly different atoms, with slightly different properties. But the model for the Hydrogen atom from the last Section suggests a way to reduce the complexity and see an underlying pattern.

Recall the formula for Hydrogen atom charge separation:

$$r_e + r_p = 32 m_p e^2 \big/ 3 m_e^2 c^2 = 5.5 \times 10^{-9}$$

The separation $r_e + r_p$ scales with the proton mass, m_p. That means the basic orbit energy of the Hydrogen atom scales with $1/m_p$. It suggests that the basic orbit energy of an arbitrary atom may scale with Z/M, where Z is the nuclear charge, and M is the nuclear mass. Therefore, scaling the measured first-order ionization potentials for all elements by the inverse factor, M/Z, may reveal some previously hidden pattern.

We find that the scaled first-order ionization potentials (called $IP_{1,Z}$) do indeed fall into a *very* regular pattern: the rise on *every* period in the Periodic Table is *exactly the same factor*, $7/2$. The consistency of the $7/2$ rise on every period suggests that the IP values are not just about individual elements; they are following a pattern that relates *all* elements, and because of the pattern, each $IP_{1,Z}$ contains information about all *other* elements. That is, $IP_{1,Z}$ embodies *population generic information*.

The breakdown that suggests itself is this: each $IP_{1,Z}$ contains a universal baseline contribution $IP_{1,1}$ about interaction between the nucleus and the population of electrons as a whole, and, for all elements beyond Hydrogen, a contribution $\Delta IP_{1,Z}$ about interactions just among the electrons by themselves.

The electron-electron contribution can be dramatic. Consider Helium: $\Delta IP_{1,2}$ is *huge*, meaning that two electrons make a very strong bond between themselves. And consider Lithium: $\Delta IP_{1,3}$ is negative, meaning that two electrons actively work together to try to exclude a third electron.

Over the periods, there is obvious detail about the electron-electron interactions. Within each period, there are obvious sub-periods keyed to the nominal angular-momentum quantum number that is being filled. Plotted on a log scale, all sub-period rises are straight lines. The slopes all appear to be rational fractions. Table 1 displays these fractions.

Table 1. Ionization Potential Rises in the Periodic Table

period	N	l	fraction	l	fraction	l	fraction	l	fraction
1	1	0	1						
2	2	0	$1/2$	1	$3/4$				
3	2	0	$1/2$	1	$3/4$				
4	3	0	$1/4$	2	$5/18$	1	$2/3$		
5	3	0	$1/4$	2	$5/18$	1	$2/3$		
6	4	0	$1/4$	3	$7/48$	2	$5/16$	1	$9/16$
7	4	0	$1/4$	3	$7/48$	2	$5/16$	1	$9/16$

Observe that a non-traditional parameter N is included in the display. For $l > 0$, this N makes it possible to write a simple formula for the fraction needed:

$$\text{fraction} = \left[(2l+1) \big/ N^2 \right]\left[(N-l)/l \right] .$$

Also, N makes it possible to characterize the lengths of all periods in the Periodic Table: the period length is always $2N^2$.

All this numerical regularity strongly suggests that there really is a reliable pattern here, and that we can reasonably seek to exploit it for calculating what events to expect in Chemical interactions. Here is the first exploitation that suggests itself: Given first-order ionization potentials of many elements, we can make reasonable estimates of the additional energy required to remove a second electron from each, and then a third, and so on.

Please note: estimating the energy to remove a second electron, or a third, and so on, is *not* the same thing as estimating the so-called 'second-order ionization potential', 'third-order ionization potential', and so on. Look up those energies, and you will see that they are very large. The implication is that the big numbers are about events that are very *violent*: ripping two, or three, or more, electrons off an atom *all at once*. This isn't what happens in ordinary every-day lab-bench Chemistry: most ionization events occur *gently*, one at a time.

Just as for the first-order ionization potentials, there is a great deal of numerical regularity in the higher-order ionization potentials. You can read about it, and view a big display Figure about it, in Chapter 1 of [1]. But the subject of the present paper does not require all this information. That is

because violent events occur only when some really heavy-duty equipment, such as a HF reactor, is involved. That is not the case for CF. The basic chemistry of water, involving only first-order ionization potentials, is enough for CF. [14]

The following development represents a simplification over that in [1]: it consolidates formulae about the one-by-one removal or addition of a number of electrons, into formulae about cumulative removal or addition of that number of electrons. This consolidation does not change end results, but it does save some computation work along the way.

The formulae all involve increments of energy identified as work W and heat H. Work is energy required to move electrons from one atom to another atom. Heat is about internal readjustments that electron populations within the atoms then make in response.

n an atom of the element with nuclear charge Z, the work to remove one electron from the nuclear field is:

$$W_{\text{removing e}_1 \text{ from the neutral atom}} = IP_{1,1}(Z / M_Z) \ .$$

This is energy that has to be supplied to get the removal process started: energy to get over a potential wall. Note the factor of Z / M. It restores *population-generic* information, IP's, to *element-specific* information.

The subsequent re-adjustment among electrons within the atom involves heat:

$$H_{\text{removing e}_1 \text{ from the neutral atom}} = (\Delta IP_{1,Z} - \Delta IP_{1,Z-1})(Z / M_Z) \ .$$

Note: for the special case $Z = 1$, $\Delta IP_{1,Z-1} = \Delta IP_{1,0}$ does not exist, and $\Delta IP_{1,Z} = \Delta IP_{1,1} = 0$. And for $Z = 2$, and $\Delta IP_{1,Z-1} = \Delta IP_{1,1} = 0$.

Thus the total energy involved in removing one electron from an atom with nuclear charge Z is:

$$(W + H)_{\text{removing e}_1 \text{ from the neutral atom}} = \left[IP_{1,1}Z + \Delta IP_{1,Z}Z - \Delta IP_{1,Z-1}(Z-1) \right] \Big/ M_Z$$

Now suppose that we remove a second electron. The work that has to be supplied because of the nuclear field is:

$$W_{\text{removing } e_2 \text{ after } e_1} = IP_{1,1}\sqrt{Z(Z-1)}\Big/M_Z \ .$$

Observe the factor of $\sqrt{Z(Z-1)}$, instead of just Z; it takes account of the difference between nuclear charge and electron count.

The heat then involved in electron readjustment is:

$$H_{\text{removing } e_2 \text{ after } e_1} = \Delta IP_{1,Z-1}(Z-1)/M_Z - \Delta IP_{1,Z-2}(Z-2)/M_Z \ .$$

The total energy involved in this step is then:

$$(W+H)_{\text{removing } e_2 \text{ after } e_1} =$$
$$IP_{1,1}\sqrt{Z(Z-1)}\Big/M_Z + \Delta IP_{1,Z-1}(Z-1)/M_Z - \Delta IP_{1,Z-2}(Z-2)/M_Z$$

Now put both steps together. The grand total energy involved in removing two electrons is then:

$$(W+H)_{\text{removing } e_1 \& e_2} = I$$
$$P_{1,1}\Big[Z+\sqrt{Z(Z-1)}\Big]\Big/M_Z + \Big[\Delta IP_{1,Z}Z - \Delta IP_{1,Z-2}(Z-2)\Big]\Big/M_Z \ .$$

Note the cancellation of $\Delta IP_{1,Z-1}(Z-1)/M_Z$ terms; such cancellations save some computation in getting to higher orders of ionization. The grand total energy involved in removing three electrons will be just:

$$(W+H)_{\text{removing } e_1, e_2, \& e_3} =$$
$$P_{1,1}\Big[Z+\sqrt{Z(Z-1)}+\sqrt{Z(Z-2)}\Big]\Big/M_Z + \Big[\Delta IP_{1,Z}Z - \Delta IP_{1,Z-3}(Z-3)\Big]\Big/M_Z \ ,$$

and so on, in that pattern.

Now let us go in the other direction, *adding* electrons. The work involved in *adding* one electron into the nuclear field of the atom with nuclear charge Z is:

$$W_{\text{adding } e_1 \text{ to neutral atom}} = -IP_{1,1}\sqrt{Z(Z+1)}\Big/M_Z \ .$$

The heat involved in then re-adjusting the electron population is:

$$H_{\text{adding e}_1\text{ to neutral atom}} = -\Delta IP_{1,Z+1}(Z+1)/M_Z + \Delta IP_{1,Z}Z/M_Z \ .$$

so that

$$(W+H)_{\text{adding e}_1\text{ to neutral atom}} = -\Big[IP_{1,1}\sqrt{Z(Z+1)} + \Delta IP_{1,Z+1}(Z+1) - \Delta IP_{1,Z}Z \Big]\Big/M_Z$$

If we add a second electron to the singly charged ion, it will require additional work:

$$W_{\text{adding e}_2\text{ after e}_1} = -IP_{1,1}\sqrt{Z(Z+2)}\Big/M_Z \ .$$

And it will cause a further heat adjustment:

$$H_{\text{adding e}_2\text{ after e}_1} = -\Delta IP_{1,Z+2}(Z+2)/M_Z + \Delta IP_{1,Z+1}(Z+1)/M_Z \ .$$

Putting both steps together, the grand total energy involved in adding two electrons is:

$$(W+H)_{\text{adding e}_1\text{ \& e}_2}$$
$$= -IP_{1,1}\Big[\sqrt{Z(Z+1)}+\sqrt{Z(Z+2)}\Big]\Big/M_Z + \Big[\Delta IP_{1,Z}Z - \Delta IP_{1,Z+2}(Z+2)\Big]\Big/M_Z.$$

Similarly, the total energy involved in adding three electrons is:

$$(W+H)_{\text{adding e}_1,\text{e}_2\text{ \& e}_3} =$$
$$-IP_{1,1}\Big[\sqrt{Z(Z+1)}+\sqrt{Z(Z+2)}+\sqrt{Z(Z+3)}\Big]\Big/M_Z + \Big[\Delta IP_{1,Z}Z - \Delta IP_{1,Z+3}(Z+3)\Big]\Big/M_Z.$$

And so it goes on, in that pattern.

The data needed for study of Water Chemistry and CF are a very small subset of the data for all the elements known, or anticipated, to exist. The full data set is available in [1]. But here we need at most the data for the first ten elements. These data, expressed in electron volts, eV, are as follows:

Hydrogen: $Z=1$, $M=1.008$, $IP_{1,1}=14.250\text{eV}$, $\Delta IP_{1,1}=0\text{eV}$

Helium: $Z = 2$, $M = 4.003$, $IP_{1,2} = 49.875\,\text{eV}$, $\Delta IP_{1,2} = 35.625\,\text{eV}$

Lithium: $Z = 3$, $M = 6.941$, $IP_{1,3} = 12.469\,\text{eV}$, $\Delta IP_{1,3} = -1.781\,\text{eV}$

Beryllium: $Z = 4$, $M = 9.012$, $IP_{1,4} = 23.327\,\text{eV}$, $\Delta IP_{1,4} = 9.077\,\text{eV}$

Boron: $Z = 5$, $M = 10.811$, $IP_{1,5} = 17.055\,\text{eV}$, $\Delta IP_{1,5} = 2.805\,\text{eV}$

Carbon: $Z = 6$, $M = 12.011$, $IP_{1,6} = 21.570\,\text{eV}$, $\Delta IP_{1,6} = 7.320\,\text{eV}$

Nitrogen: $Z = 7$, $M = 14.007$, $IP_{1,7} = 27.281\,\text{eV}$, $\Delta IP_{1,7} = 13.031\,\text{eV}$

Oxygen: $Z = 8$, $M = 15.999$, $IP_{1,8} = 27.281\,\text{eV}$, $\Delta IP_{1,8} = 13.031\,\text{eV}$

Fluorine: $Z = 9$, $M = 18.998$, $IP_{1,9} = 34.504\,\text{eV}$, $\Delta IP_{1,9} = 20.254\,\text{eV}$

Neon: $Z = 10$, $M = 20.180$, $IP_{1,10} = 43.641\,\text{eV}$, $\Delta IP_{1,10} = 29.391\,\text{eV}$

THE BASICS OF WATER CHEMISTRY

Ordinary Liquid Water

Ordinary liquid water illustrates an important point: sometimes widely held, but not really examined, assumptions can turn out to be questionable. For example, what is the likely ionic configuration of a water molecule?

One possible ionic configuration for water is $2\,H^+ + O^{2-}$, and most people believe this is what ordinary water is. Let us now check that belief. The transition $H \rightarrow H^+$ takes:

$$IP_{1,1} / M_1 = 14.250 / 1.008 = 14.1369\,\text{eV}\,.$$

So $2H^+$ takes $2 \times 14.1369 = 28.2738\,\text{eV}$. The transition $O \rightarrow O^{2-}$ takes:

$$-IP_{1,1}\left[\sqrt{Z(Z+1)}+\sqrt{Z(Z+2)}\right]\Big/M_Z+\left[\Delta IP_{1,Z}Z-\Delta IP_{1,Z+2}(Z+2)\right]\Big/M_Z$$

$$=-14.250\left[\sqrt{8\times9}+\sqrt{8\times10}\right]\Big/15.999+\left[\Delta IP_{1,8}\times8-\Delta IP_{1,10}\times10\right]\Big/15.999$$

$$=-14.250\left[\sqrt{72}+\sqrt{80}\right]\Big/15.999+\left[13.031\times8-29.391\times10\right]\Big/15.999$$

$$=-14.250[8.4853+8.9443]/15.999+[104.248-293.910]/15.999$$

$$=-14.250[17.4296]/15.999+[-189.662]/15.999$$

$$=-15.5242-11.8546=-27.3788\text{eV}\quad.$$

So the creation of a water molecule in the ionic configuration $2\,\text{H}^+ + \text{O}^{2-}$ takes altogether $28.2738 - 27.3788 = 0.8950\text{eV}$. This is a positive energy requirement, and it implies that some external assistance is needed. So normal water may *not* be in this ionic configuration after all!

But another possibility is readily at hand. The ionic configuration for normal water could be $2\,\text{H}^- + \text{O}^{2+}$. The transition $\text{H}\rightarrow\text{H}^-$ takes:

$$-IP_{1,1}\sqrt{Z(Z+1)}\Big/M_Z-\Delta IP_{1,Z+1}(Z+1)/M_Z+\Delta IP_{1,Z}Z/M_Z$$

$$=-IP_{1,1}\sqrt{1\times2}\Big/M_1-\Delta IP_{1,2}\times2/M_1+\Delta IP_{1,1}\times1/M_1$$

$$=-14.250\times1.4142/1.008-35.625\times2/1.008+0\times1/1.008$$

$$=-19.9924-70.6845+0=-90.6769\text{eV}.$$

So $2\,\text{H}^-$ takes $2\times(-90.6769)=-181.3538\text{eV}$. Please pause here and take notice: this is a *huge* negative energy. It reflects the fact that electrons *really* like to make pairs. They do it everywhere in Chemistry. Their propensity to do it motivated the invention of the so-called 'spin' quantum number. Without spin, many electrons in atoms would be violating the 'Pauli Exclusion Principle', which says that only one electron can be in any particular quantum state.

Electron pairs are famous in Condensed Matter Physics too, under the name 'Cooper pairs'.

Now proceeding, the transition $\text{O}\rightarrow\text{O}^{2+}$ takes:

$$IP_{1,1}\left[Z+\sqrt{Z(Z-1)}\right]\Big/M_Z+\left[\Delta IP_{1,Z}Z-\Delta IP_{1,Z-2}(Z-2)\right]\Big/M_Z$$

$$=IP_{1,1}\left[8+\sqrt{8\times7}\right]\Big/M_8+\left[\Delta IP_{1,8}\times8-\Delta IP_{1,6}\times6\right]\Big/M_8$$

$$=14.250\left[8+\sqrt{56}\right]\Big/15.999+\left[13.031\times8-7.320\times6\right]/15.999$$
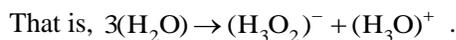
$$=13.7907+3.7707=17.5614\,\text{eV}\,.$$

So the creation of the water molecule in the ionic configuration $2\,H^-+O^{2+}$ takes altogether $2\times(-90.6769)+17.5614=-181.3538+17.5614=-163.7924\,\text{eV}$. This is a solidly negative energy. It means that ordinary water is overwhelmingly in the second candidate ionic configuration, $2\,H^-+O^{2+}$. This *is* normal water.

However, in situations where more than one version of anything can exist, they generally all *do* exist, in proportions determined by their so-called Boltzmann factors, $\exp(-E/kT)$, where E is energy, k is Boltzmann's constant, and T is absolute temperature. Boltzmann factors are the evidence of entropy maximization at work. So ordinary liquid water will contain a tiny, tiny fraction of the first ionic configuration, $2H^++O^{2-}$.

EXCLUSION ZONE WATER

Prof. Gerald Pollack wrote the definitive book [15] on the subject of water in the newly discovered state characteristic of the surface of all water. This state is called Exclusion Zone Water, or EZ water. The name comes from the fact that EZ water tends to expel positive ions. It does this because it forms nested layers of hexagonal arrays of negative ions, each layer offset to the next, thus leaving no space for unused positive ions.

The following chemical transition is proposed for the creation of EZ water:

3 normal water molecules $\rightarrow$ 1 EZ water ion + 1 hydronium ion

That is, $3(H_2O)\rightarrow(H_3O_2)^-+(H_3O)^+$.

Observe the notation: parentheses are used to avoid any prejudice about what charge the individual atoms within any chemical species, whether neutral

molecule or charged ion or radical, may carry. We want to consider all possible, or at least all plausible, ionic configurations of every chemical species mentioned.

The EZ water ion $(H_3O_2)^-$ could have the ionic configuration $3\,H^+ + 2\,O^{2-}$. The $3\,H^+$ takes total energy $3 \times IP_{1,1} / M_z = 3 \times 14.250 / 1.008 = 42.4107\,eV$. The $2\,O^{2-}$ takes total energy $2 \times (-27.3788) = -54.7576\,eV$. So the ionic configuration $3\,H^+ + 2\,O^{2-}$ takes $42.4107 - 54.7576 = -12.3469\,eV$. This is a negative energy, so this ionic configuration certainly can occur.

But there is another possibility too. The EZ water ion $(H_3O_2)^-$ could have the ionic configuration $3\,H^- + 2\,O^+$. An H^- takes $-90.6769\,eV$, so $3\,H^-$ takes $3 \times (-90.6769) = -272.0307\,eV$. An O^+ takes energy:

$$(IP_{1,1} \times 8 + \Delta IP_{1,8} \times 8 - \Delta IP_{1,7} \times 7)/M_8$$
$$= (14.250 \times 8 + 13.031 \times 8 - 13.031 \times 7)/15.999$$
$$= (114.000 + 104.248 - 91.217)/15.999 = 7.9399\,eV.$$

So $2\,O^+$ takes $2 \times 7.9399 = 15.8798\,eV$. Thus the ionic configuration $3\,H^- + 2\,O^+$ takes $-272.0307 + 15.8798 = -256.1509\,eV$. This energy is much more negative than that of the first candidate ionic configuration for EZ water, $3\,H^+ + 2\,O^{2-}$. This fact means EZ water is nearly always in this second candidate ionic configuration, $3\,H^- + 2\,O^+$.

As for the hydronium ion $(H_3O)^+$, it could have the ionic configuration $3\,H^+ + O^{2-}$. This, I believe, is what most people would expect, given that most people would have expected (wrongly, as it turned out) that normal water had the ionic configuration $2\,H^+ + O^{2-}$. From the study of regular water, we know the candidate $3\,H^+$ would take $3 \times 14.250 = 42.4107\,eV$, and the candidate O^{2-} would take $-27.3788\,eV$, so the candidate ionic configuration $3\,H^+ + O^{2-}$ for hydronium would take $42.4107 - 27.3788 = 15.0319\,eV$. This

is a solidly positive energy, *and so the ionic configuration* $3\,H^+ + O^{2-}$ *for the hydronium ion* $(H_3O)^+$ *occurs only very infrequently.*

But, as it was with the EZ water ion, there is another possibility for the hydronium ion. The $(H_3O)^+$ could have the ionic configuration $3\,H^- + O^{4+}$. The $3\,H^-$ would take $3 \times (-90.6769) = -272.0307\,eV$, and the O^{4+} would take:

$$IP_{1,1}\left[Z + \sqrt{Z(Z-1)} + \sqrt{Z(Z-2)} + \sqrt{Z(Z-3)}\right]\Big/M_Z + \left[\Delta IP_{1,Z}Z - \Delta IP_{1,Z-4}(Z-4)\right]\Big/M_Z$$

$$= 14.250 \times \left[8 + \sqrt{56} + \sqrt{48} + \sqrt{40}\right]\Big/15.999 + \left[13.031 \times 8 - 9.077 \times 4\right]/15.999$$

$$= 14.250 \times 28.7361\,/\,15.999 + 67.940\,/\,15.999 = 25.5947 + 4.2465 = 29.8411\,eV.$$

So the second candidate ionic configuration $3\,H^- + O^{4+}$ for the hydronium ion $(H_3O)^+$ would take $-272.0307 + 29.8411 = -242.1895\,eV$. This is a robustly negative energy. It explains why the reaction product that accompanies EZ water *is* a hydronium ion, rather than a naked proton ($14.1369\,eV$) next to a normal water molecule ($-163.7925\,eV$), which would add up to $-149.6556\,eV$, which, though negative, is not anywhere near as robustly so.

Together, the EZ water ion and the hydronium ion take $-268.6204 - 242.1895 = -510.8099\,eV$. Compare this to three normal water molecules at $3 \times (-163.7925) = -491.3775\,eV$. The EZ water ion together with the hydronium ion is favored over the three normal water molecules. That means nature stands poised to make EZ water ions and hydronium ions whenever it has the opportunity to do so. Generally speaking, the opportunity involves a surface and a little bit of incident energy, such as that provided by sunlight.

The other known states of water are Water Ice and Water Vapor, both of great interest to be sure, but probably not directly involved in CF.

THE CF APPLICATION

The next question in the present study is this: what if any of the molecules or ions discussed here had been created with some heavy nuclei included?

That is, what happens to the ionic configurations if one Hydrogen atom H is replaced with a Deuterium D , or even a Tritium T ?

where an H^+ ion takes:

$$IP_{1,1} \times 1 / M_H = IP_{1,1} \times 1/1 = 14.2500 \text{ eV} ,$$

a D^+ ion takes:

$$IP_{1,1} \times 1 / M_D = IP_{1,1} \times 1/2 = 14.250/2 = 7.1250 \text{ eV} ,$$

and a T^+ ion takes:

$$IP_{1,1} \times 1 / M_T = IP_{1,1} \times 1/3 = 14.250/3 = 4.7500 \text{ eV} .$$

The message is: heavier positive ions are easier to make.
Similarly, where an H^- ion takes:

$$-IP_{1,1} \sqrt{1 \times 2} / M_H - \Delta IP_{1,2} \times 2 / M_H =$$
$$-14.250 \times 1.4142 / 1 - 35.6250 \times 2 / 1$$
$$= -20.1495 - 71.2500 = -91.3995 \text{ eV}$$

a D^- instead takes:

$$-IP_{1,1} \sqrt{1 \times 2} / M_D - \Delta IP_{1,2} \times 2 / M_D =$$
$$-14.2500 \times 1.4142 / 2 - 35.6250 \times 2 / 2 =$$
$$-10.0762 - 35.6250 = -45.7012 \text{ eV}$$

and a T^- takes:

$$-IP_{1,1} \sqrt{1 \times 2} / M_T - \Delta IP_{1,2} \times 2 / M_T =$$
$$-14.2500 \times 1.4142 / 3 - 35.6250 \times 2 / 3 =$$
$$-6.7175 - 23.7500 = -30.4675 \text{ eV}$$

The message is: heavier negative ions are less rewarding to make.

This is all to the good, inasmuch as CF requires naked nuclei, i.e., positive ions, as reactants. So can we get such reactants from water? In the case of normal water, we now know that the usual ionic configuration must be $2\,H^- + O^{2+}$, and not the previously presumed $2\,H^+ + O^{2-}$. The latter has a positive energy, and so does not occur abundantly.

But what about $D^+ + H^+ + O^{2-}$? We have above that D^+ takes 7.1250 eV, H^+ takes 14.2500 eV, and from the earlier analysis of ordinary water, we know O^{2-} takes -27.3788 eV. So $D^+ + H^+ + O^{2-}$ will take

$$7.1250 + 14.2500 - 27.3788 = -6.0038\ \text{eV}\ .$$

So in heavy water, this ionic configuration will occur naturally too.

In the case of EZ water, we know that, along with the more favored ionic configuration $3\,H^- + 2\,O^+$, the less favored $3\,H^+ + 2\,O^{2-}$ *can* exist naturally, and so probably *does* exist. So what about $D^+ + 2\,H^+ + 2\,O^{2-}$? That will take

$$7.1250 + 2 \times 14.2500 + 2 \times (-27.3788) =$$
$$7.1250 + 28.5000 - 54.7576 = -19.1326\ \text{eV}\ .$$

Compare this to $3\,H^+ + 2\,O^{2-}$ with pure Hydrogen, at

$$3 \times 14.250 / 1 - 54.7576 = 42.750 - 54.7576 = -12.0076\ \text{eV}\ .$$

The EZ ion with a Deuterium ion included is at *lower energy*, and so forms *more easily* than the one without Deuterium. This EZ ion will be preferred. So the action of forming EZ water can tend to concentrate the kind of reactant that CF requires.

But we should also look at the other ionic configuration for a deuterated EZ ion, $D^- + 2\,H^- + 2\,O^+$. The D^- takes

$$-IP_{1,1}\sqrt{Z(Z+1)}\Big/M_Z - \Delta IP_{1,Z+1}(Z+1)/M_Z + \Delta IP_{1,Z}Z/M_Z$$
$$= -IP_{1,1}\sqrt{1 \times 2}\Big/2 - \Delta IP_{1,2}\,2/2 + \Delta IP_{1,1}/2$$
$$= -14.250 \times 1.4142/2 - 35.625 + 0 = -25.5486\ \text{eV}.$$

Above we found that pure H^- takes $-20.1495 - 71.2500 = -91.3995$ eV, so $2\,H^-$ takes $2 \times (-91.3995) = -182.7990$ eV. The $2\,O^+$ takes $2 \times 7.9399 = 15.8798$ eV. So altogether, the ionic configuration $D^- + 2\,H^- + 2\,O^{2+}$ takes:

$$-25.5486 - 182.7990 + 15.8798 = -192.4678\,\text{eV}\ .$$

This is a solidly negative energy, so this ionic configuration would certainly dominate the EZ – at least until such time a source of energy is introduced. That source would be the electrodes that are inserted to start the CF attempt. The process is analogous to the process for making Brown's gas: electrochemistry. In the CF case, when the electrodes are introduced they supply some energy, and the EZ ions can then go to their less favored, higher-energy, ionic configuration, wherein the deuterons are exposed, and CF can then ensue. This EZ story appears to play into a peculiar fact about CF that is presently known, but not well understood: CF generally goes better if the reactant heavy water is left to rest quietly for a long time before being used. This rest period was described as 'incubation' in a talk given by Senior VP Frank Gordon at Global Energy Corporation, Annandale, VA, speaking at a Conference of the Natural Philosophy Alliance in Long Beach, CA, in 2010.

Since then, I have come to think that three important processes may characterize 'incubation': 1) Gravity sends water molecules with heavier isotopes to lower levels in the water container, thus concentrating them more at locations lower in the water container; 2) The water container has internal surfaces, and they encourages EZ formation at those surfaces. There, the exclusion property of the EZ water further concentrates the positive ions that contain the heavy isotopes. 3) The stage is then set, but nothing dramatic can happen until electrodes are inserted, and a little energy is supplied to trigger the CF process. Where should the electrodes go? We may want to avoid, *or* to seek, the exclusion zones at the top, bottom, and sides of the container. We certainly want to reach positive ions of heavier mass, so we should put electrodes in the lower part of the water container. Should the current go horizontal or vertical, and if vertical, which way? We need experimental investigation of all these questions.

CONCLUSION

CF is an Engineering Problem, and it demands Engineering Analyses. This paper offers some suggestions as to what the engineering analyses should be like.

1) Engineering is somewhat like Applied Mathematics. The DSB protocol is to Applied Mathematics what First Aid is to Medicine: Step 1. In many cases, it solves the whole problem at hand. Only when it does *not* seem to solve the whole problem should the practitioner turn to anything else. So I believe the Engineering Analysis of CF should definitely use the DSB Protocol.

2) Engineering is usually *not* like Pure Mathematics: engineers do not generally involve Postulates in Engineering Analyses. What they do instead is list their Assumptions. The word 'Assumption' is much milder than the word 'Postulate'; it allows for a possible update, such as may occur in a typical engineering design review. So I believe engineers should not feel bound by any pre-established Postulate, such as the Second Postulate of SRT.

I am today hopeful that we can solve many of the engineering problems presented by CF. In so doing, we might also help eliminate the many Paradoxes that come into Physics along with the Second Postulate of SRT. The present situation is *itself* quite paradoxical: It is generally considered *radical* to do the actually *traditional* thing; i.e., apply the DSB Protocol, and it is generally considered *traditional* to do the originally *radical* thing; i.e., embrace the Second Postulate!

REFERENCES

[1]   C.K. Whitney, *Algebraic Chemistry, Applications and Origins*, Nova Science Publishers, Hauppauge, NY, USA (2013).

[2]   A. Einstein, "On a Heuristic Point of View Concerning the Production and Transformation of Light", Annalen der Physik 17, 132-148, 1905, reprinted in *The Collected Papers of Albert Einstein*, Vol. 2, pp. 86 – 103, Princeton University Press, (1989).

[3]   A. Einstein, "On the Electrodynamics of Moving Bodies", Annalen der Physik 17, 891-921, 1905, reprinted in *The Collected Papers of Albert Einstein*, Vol. 2, pp. 140 – 173, Princeton University Press (1989).

[4]   R.P. Feynman, QED, *The Strange Theory of Light and Matter*, Princeton University Press, Princeton, NJ (1985).

[5]   L. Brilloin, *Science and Information Theory*, Second Ed., Dover Publications, Mineola, NY (1962).

[6]   http://www.britannica.com/EBchecked/topic/194901/Euclidean-geometry.

[7]   J.D. Jackson, *Classical Electrodynamics*, Sect. 14.1, Second Ed., John Wiley & Sons, New York, (1962, 1975).

[8]   W. Ritz, *Recherches critiques sur l'Électrodynamique Générale*, *Annales de Chimie et de Physique* 13: 145–275 (1908).

[9]   V. Kuligin, M. Korneva, G. Kuligina, "Lorentz Transformation and Image Distortion", Galilean Electrodynamics 26, Special Issue 2, Fall (2015) (forthcoming).

[10]   A. Liènard, "Champ Electrique et Magnétique produit par une Charge Electrique Concentrée en un Point et Animée d'un Movement Quelconque", *L'Eclairage Electrique vol. XVI* (27) 5-14, (28) 53-59, (29) 106-112 (1898).

[11]   E. Wiechert, "Elektrodynamische Elementargesetze", *Archives Néerlandesises des Sciences Exactes et Naturelles série II, Tome IV* (1901).

[12]   L.H. Thomas, "The Kinematics of an Electron with an Axis", Phil. Mag. Series 7, 3 (13) 1 – 22 (1927).

[13]   J.P. Wesley, *Classical Quantum Theory*, Capter 7, Benjamin Wesley, 78176 Blumberg, Germany, 1996.

[14]   C. Whitney, "EZ Water: A New Quantitative Approach Applied – Many Numerical Results Obtained", Water 5, 105-120 (2013).

[15]   G. Pollack, *The Fourth Phase of Water: Beyond Solid, Liquid, and Vapor*, Ebner & Sons, Seattle WA, USA (2013).