# NONISOTROPIC SPATIOTEMPORAL CHAOTIC VIBRATION OF THE WAVE EQUATION DUE TO MIXING ENERGY TRANSPORT AND A VAN DER POL BOUNDARY CONDITION 

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#### Abstract

The imbalance of the boundary energy flow due to energy injection at one end and a nonlinear van der Pol boundary condition at the other end of the spatial one-dimensional interval can cause chaotic vibration of the linear wave equation [Chen et al., 1998b, 1998c]. However, such chaotic vibration is isotropic with respect to space and time because the two associated families of characteristics both propagate with the same speed and, thus, the "strength of chaos" is the same along both $x$ and $t$ directions. In this paper, we show that by including a mixed partial derivative linear energy transport term in the wave equation, nonlinearity in the van der Pol boundary condition can also cause chaotic vibration (without energy injection from the other end). Two new families of characteristics now travel with different speeds, leading to strong mixing of waves and nonisotropic spatiotemporal chaos. Parameter range for the route to chaos, including period-doubling, homoclinic orbits and Cantor-like invariant sets is classified. Numerical simulations of chaotic space-time profiles are also illustrated.


Keywords: Nonisotropic spatiotemporal chaos; wave equation; van der Pol boundary condition.

## 1. Introduction and Model Formulation

Earlier, in a series of papers [Chen et al., 1998a1998d], we studied nonlinear vibration of the onedimensional (1D) wave equation on a bounded interval with a van der Pol boundary condition. Chaos occurs, for example, when there is energy injection either from the other end of the interval [Chen et al., 1998b] or at the middle-of-
the-span point [Chen et al., 1998d]. Related studies have been done in [Sharkovsky et al., 1993, 1994] and [Shimura et al., 1967], for example. A major reason that has motivated us to study such problems is that the 1 D wave equation is a rather basic distributed parameter model for mechanical vibration, such as vibrating strings and acoustic wave propagation. With given or designed nonlinear boundary conditions/controllers,

[^0]the solution is amenable to a thoroughly rigorous mathematical analysis, revealing immensely rich dynamical behavior suitable for possible future technological applications. On the one hand, we have tried to explore the useful, microscosmic nature of spatiotemporal chaos in 1D wave propagation, in particular, in acoustics; on the other hand, the knowledge gained will improve our understanding of macrocosmic chaos in complex, general distributed parameter systems, which will be the ultimate objective for investigation.

Elsewhere, in engineering, as far as acoustic chaos is concerned, there has been a sharp surge of interest in recent years. A good number of papers [Herzel, 1993; Holzfuss \& Lauterborn, 1989; Idogawa et al., 1993; Katz, 1996; Lauterborn et al., 1996; Lauterborn \& Cramer, 1981; Lauterborn \& Holzfuss, 1986; Mcintyre et al., 1983; Mende et al., 1990; Shigesada et al., 1979; Smith et al., 1982; Stinecke \& Herzel, 1995; Swift, 1988, 1995; Tufilaro, 1989; Yazaki, 1993], have been published. We may mention, for example, the following research and applications:
(i) In [Lauterborn et al., 1996], the collaborative German and Russian scientists studied chaotic motions in acoustic cavitation, which understandably is important for the detection of a submarine's underwater acoustic signature;
(ii) the development of musical instruments that play new types of sounds generated by chaotic acoustics, along with the corresponding musical notes to be composed by musicians [Idogawa et al., 1993; Swift, 1988, 1995];
(iii) study of sounds made by the vocal cords in the speech production process has revealed bifurcation and chaotic dynamics that could be used for personal (voice) identification and health diagnosis (for newborn infants from their cries [Mende et al., 1990]);
(iv) the design of thermoacoustic oscillators as possible musical instruments playing both ordinary and chaotic musical notes, and cooling/ heating thermoacoustic engines whose control is achieved by boundary heating/cooling [Swift, 1988, 1995].

Most of the above-mentioned references emphasize experimental and/or modeling work with realworld applications, but did not provide rigorous mathematical analysis. Nevertheless, some of their physical motivations are quite inspiring; see the discussion below.

Return to our work [Chen et al., 1998a-1998d]. The study of the chaotic vibration of the 1D wave equation therein, although spatiotemporal in nature, is essentially isotropic with respect to space and time. There are at least two reasons for this: first, the wave equation $w_{x x}-w_{t t}=0$ is invariant under the changes of variables

$$
\begin{equation*}
(x, t) \rightarrow(-x, t),(x, t) \rightarrow(x,-t),(x, t) \rightarrow(t, x) \tag{1}
\end{equation*}
$$

and, thus, the symmetry structure between $x$ and $\pm t$ of the equation itself is extremely strong. (Indeed, if there were no boundary conditions, then $x$ and $\pm t$ may be regarded as "identical".) Second, the two families of characteristics belong to the wave equation both travel with the same speed, again making $x$ and $\pm t$ indiscernible and, thus, we may say that "the strength of chaos" is the same in both the $x$ and $t$ directions. The spatiotemporalisotropic nature of the solutions studied in [Chen et al., 1998a-1998d] is evident in explicit representations like [Chen et al., 1998b, p. 425, (13), (14)], and in the many graphics of solution profiles displayed in [Chen et al., 1998b, 1998c]. For an immediate comparison, see the graphics in Sec. 6 below.

Obviously, for nonisotropic spatiotemporal chaos to happen, a key element is to have a medium which is not spatially homogeneous. Thus, for the wave equation, we need something else in the equation that is not invariant with respect to the change of variables (1). One might attempt to add a variable coefficient term (say, $a(x) w(x, t)$ ) to the wave equation. However, this makes the mathematical treatment rather difficult. Here we have found a sufficiently simple medium in which waves traveling to the left and the right have different speeds. The model equation in this paper is the following hyperbolic equation

$$
\begin{gather*}
w_{x x}(x, t)-\nu w_{x t}(x, t)-w_{t t}(x, t)=0 \\
0<x<1, \quad t>0 \tag{2}
\end{gather*}
$$

where $\nu$ is a positive constant and where, for the mere sake of convenience, we have set the coefficient for the $-w_{t t}$ term in (7) to be 1 as well, because the mathematical analysis in this paper goes through with only straightforward modifications as long as $-w_{t t}$ is multiplied by a positive constant. Equation (2) differs from the standard wave equation $w_{x x}-w_{t t}=0$ by the presence of the "mixed term" $-\nu w_{x t}$. Let us now make precise the physical
origin of this term. Consider acoustic waves traveling in a pipe filled with a gas moving with speed $c: 0<c<1$ to the left, such as air blown into a clarinet in the aforementioned papers [Idogawa et al., 1993; Swift, 1988, 1995] on chaotic musical instruments. Assume that an observer travels with the gas, measures the acoustic pressure field and finds that it satisfies the wave equation

$$
\begin{equation*}
\frac{\partial^{2} w(x, t)}{\partial x^{2}}-\frac{\partial^{2} w(x, t)}{\partial t^{2}}=0 . \tag{3}
\end{equation*}
$$

A stationary observer is also making measurements of the same acoustic wave propagation. His coordinate system is therefore

$$
\begin{equation*}
x^{\prime}=x+c t, \quad t^{\prime}=t \tag{4}
\end{equation*}
$$

He will find that the equation for the acoustic pressure field is necessarily

$$
\begin{align*}
& \left(1-c^{2}\right) \frac{\partial^{2} w\left(x^{\prime}, t^{\prime}\right)}{\partial x^{\prime 2}}-2 c \frac{\partial^{2} w\left(x^{\prime}, t^{\prime}\right)}{\partial x^{\prime} \partial t^{\prime}} \\
& -\frac{\partial^{2} w\left(x^{\prime}, t^{\prime}\right)}{\partial t^{\prime 2}}=0 \tag{5}
\end{align*}
$$

i.e. the equation derived from (3) by the change of variables (4). If we normalize the coefficient of $\partial^{2} w / \partial x^{\prime 2}$ in (5) to 1 by another change of variables

$$
\begin{equation*}
x=\frac{1}{\sqrt{1-c^{2}}} x^{\prime}, \quad t=t^{\prime} \tag{6}
\end{equation*}
$$

we obtain (2), with $\nu \equiv 2 c / \sqrt{1-c^{2}}$.
Assume that the boundary conditions (B.C.) associated with (2) are homogeneous Dirichlet:

$$
\begin{equation*}
w(0, t)=0, \quad w(1, t)=0, \quad t>0 \tag{7}
\end{equation*}
$$

Also, let

$$
\begin{gather*}
w(x, 0)=w_{0}(x), \quad w_{t}(x, 0)=w_{1}(x)  \tag{8}\\
0<x<1
\end{gather*}
$$

be two given initial conditions (I.C.) that are sufficiently smooth. The energy of vibration of the system at time $t$ is defined to be

$$
E(t)=\frac{1}{2} \int_{0}^{1}\left[w_{x}^{2}(x, t)+w_{t}^{2}(x, t)\right] d x
$$

(This is not the original energy seminorm associated with (3) or (5) according to the changes of variables (4) or (6). However, it is equivalent to
that.) Differentiating:

$$
\begin{align*}
\frac{d}{d t} E(t) & =\int_{0}^{1}\left[w_{x} w_{x t}+w_{t} w_{t t}\right] d x \\
& =\int_{0}^{1} \frac{\partial}{\partial x}\left[w_{x} w_{t}-\frac{\nu}{2} w_{t}^{2}\right] d x \\
& =\left.\left[w_{x}(x, t) w_{t}(x, t)-\frac{\nu}{2} w_{t}^{2}(x, t)\right]\right|_{x=0} ^{x=1} \\
& =0 \tag{9}
\end{align*}
$$

and, therefore, we see that the presence of the term $-\nu w_{x t}$ does not change the total energy of the system, i.e. energy is conserved all the time. However, there is no question that the term $-\nu w_{x t}$ is capable of energy transport within the system. Indeed, as we can expect from the treatment below (cf. (17)(24)), the term $-\nu w_{x t}$ will change the speeds of wave propagation by way of which energy transport is achieved. For this reason, we give the term $-\nu w_{x t}$ a moniker LMET: linear mixing energy transport. Note that Eq. (2) is no longer invariant under the transformation (1).

Remark 1.1. There is a certain similarity between the LMET $-\nu w_{x t}$ and the convection term in unsteady fluid motions. Consider the 3D incompressible flow satisfying the Navier-Stokes equation without external forcing:

$$
\left\{\begin{align*}
\frac{\partial}{\partial t} \mathbf{u}+\mathbf{u} \cdot \nabla \mathbf{u}-\mu \Delta \mathbf{u} &  \tag{10}\\
\quad+\nabla p=0, & \text { on } \Omega, \text { for } t>0, \\
\nabla \cdot \mathbf{u}=0 & \text { on } \Omega \\
\left.\mathbf{u}\right|_{\partial \Omega=\mathbf{0},} & \\
\mathbf{u}(x, 0)=\mathbf{u}_{0}(x), & x \in \Omega,
\end{align*}\right.
$$

where $\mathbf{u}=\mathbf{u}(x, t)=\left(u_{1}(x, t), u_{2}(x, t), u_{3}(x, t)\right)$ is the velocity function at point $x=\left(x_{1}, x_{2}, x_{3}\right) \in$ $\Omega \subset \mathbb{R}^{3}$ at time $t$, on a bounded domain $\Omega$ with boundary $\partial \Omega ; p=p(x, t)$ is the pressure function and $\mu>0$ is a constant related to Reynold's number. The terms $\mathbf{u} \cdot \nabla \mathbf{u}$ and $-\mu \Delta \mathbf{u}$ in (10) ${ }_{1}$ are called, respectively, the convection and diffusion terms. The total energy of the fluid at time $t$ is

$$
E(t) \equiv \frac{1}{2} \int_{\Omega}|\mathbf{u}(x, t)|^{2} d x, \quad\left(d x=d x_{1} d x_{2} d x_{3}\right) .
$$

Assume that the solution is sufficiently smooth. Then the rate of change of energy is

$$
\begin{align*}
\frac{d}{d t} E(t)= & \int_{\Omega} \mathbf{u} \cdot \frac{\partial \mathbf{u}}{\partial t} d x \\
= & \int_{\Omega} \mathbf{u} \cdot[-\mathbf{u} \cdot \nabla \mathbf{u}+\mu \Delta \mathbf{u}-\nabla p] d x \\
& \text { (integration by parts } \Rightarrow) \\
= & \frac{1}{2}\left\{-\int_{\partial \Omega}(\mathbf{u} \cdot \mathbf{n})|\mathbf{u}|^{2} d \sigma+\int_{\Omega}(\nabla \cdot \mathbf{u})|\mathbf{u}|^{2} d x\right\} \\
& -\mu \int_{\Omega}|\nabla \mathbf{u}|^{2} d x+\mu \int_{\partial \Omega} \frac{\partial \mathbf{u}}{\partial n} \cdot \mathbf{u} d \sigma \\
& +\int_{\partial \Omega}(\mathbf{u} \cdot \mathbf{n}) p d \sigma-\int_{\Omega}(\nabla \cdot \mathbf{u}) p d x \tag{11}
\end{align*}
$$

where $\mathbf{n}$ is the unit outward normal on $\partial \Omega$. The terms inside $\{\ldots\}$ above, contributed by the convection term $\mathbf{u} \cdot \nabla \mathbf{u}$, are equal to zero because $\left.\mathbf{u}\right|_{\partial \Omega}=0$ and $\nabla \cdot \mathbf{u}=0$ on $\Omega$. Therefore convection"transports energy around in $\Omega$ ", but does not contribute to the growth or decay of the total energy. (Nevertheless, the final outcome of (11) is

$$
\frac{d}{d t} E(t)=-\mu \int_{\Omega}|\nabla \mathbf{u}|^{2} d x \leq 0
$$

and thus, there is net energy loss due to diffusion, but not due to convection.)

The similarity between the LMET $-\nu w_{x t}$ and the convection term $\mathbf{u} \cdot \nabla \mathbf{u}$ is thence obvious as far as energy transport and conservation are concerned. This is an additional physical interpretation and motivation to our discussion in (2)-(9) above.

Remark 1.2. If the boundary condition $(10)_{3}$ is changed from stationary (i.e. $\left.\mathbf{u}\right|_{\partial \Omega}=0$, the RHS of $(10)_{3}$ ) to no-slip (i.e. $\left.\mathbf{u}(x, t)\right|_{\partial \Omega}=\mathbf{U}(x, t)$, where $\mathbf{U}(x, t)$ is the velocity of the moving boundary), then the convection term $\mathbf{u} \cdot \nabla \mathbf{u}$ in (10) ${ }_{1}$ will no longer maintain a neutral role in the growth or decay of energy of fluid motion because the first term on the RHS of (11) now becomes

$$
-\frac{1}{2} \int_{\partial \Omega}(\mathbf{u} \cdot \mathbf{n})|\mathbf{u}|^{2} d \sigma=-\frac{1}{2} \int_{\partial \Omega}(\mathbf{U} \cdot \mathbf{n})|\mathbf{u}|^{2} d \sigma
$$

which is generally indefinite in sign.
Return to the study of the wave equation. We now consider the following system of initialboundary value problem

$$
\begin{cases}w_{x x}(x, t)-\nu w_{x t}(x, t)-w_{t t}(x, t)=0, & 0<x<1, \quad t>0  \tag{12}\\ w_{x}(0, t)=0, & t>0 \\ w_{x}(1, t)=\alpha w_{t}(1, t)-\beta w_{t}^{3}(1, t) ; & \alpha, \beta>0 ; \quad t>0 \\ w(x, 0)=w_{0}(x), w_{t}(x, 0)=w_{1}(x), & 0<x<1\end{cases}
$$

In comparing the above with (2)-(8), we note that the two boundary conditions have been changed: $(12)_{2}$ is the homogeneous Neumann condition instead of the homogeneous Dirichlet (7) $)_{1}$; $(12)_{3}$ is the self-exciting (or self-regulating) van der Pol boundary condition as we have studied in [Chen et al., 1998a-1998d], which is analogous to the van der Pol ODE; see [Stoker, 1950, Chap. V.A]. Equation $(12)_{3}$ is the only place where nonlinearity makes its appearance. This is significant. It is also noteworthy that we need the Neumann (i.e. $(12)_{2}$ ) condition instead of the Dirichlet (i.e. $\left.(7)_{1}\right)$ in order for chaos to happen; see the explanation in Sec. 7. For now we only require $\alpha$ and $\beta$ to be positive constants. As we make progress we will specify the range of $\alpha$ so that the "hysteresis" situation as in [Chen et al., 1998c] can be avoided and, therefore, uniqueness of solutions of (12) is ensured.

Once again, let us examine the rate of change of energy for (12):

$$
\begin{align*}
\frac{d}{d t} E(t)= & \frac{d}{d t}\left[\frac{1}{2} \int_{0}^{1}\left(w_{x}^{2}+w_{t}^{2}\right) d x\right] \\
& (\text { integration by parts } \Rightarrow) \\
= & {\left.\left[w_{x}(x, t) w_{t}(x, t)-\frac{\nu}{2} w_{t}^{2}(x, t)\right]\right|_{x=0} ^{x=1} } \\
= & T_{1}+T_{2} \tag{13}
\end{align*}
$$

where

$$
\begin{gather*}
T_{1} \equiv \frac{\nu}{2} w_{t}^{2}(0, t) \\
T_{2} \equiv w_{t}^{2}(1, t)\left[\left(\alpha-\frac{\nu}{2}\right)-\beta w_{t}^{2}(1, t)\right] \tag{14}
\end{gather*}
$$

Remark 1.3. The positivity or negativity of $T_{1}$ and $T_{2}$ in (14) signifies the following:
(i) $T_{1} \geq 0$ if $\nu>0$, i.e. there is energy injection into the system indirectly through the LMET term $-\nu w_{x t}$. This $T_{1}$ term would have disappeared if the homogeneous Dirichlet condition $w(0, t)=0$ were imposed at $x=0 ; \mathrm{cf}$. Remark 1.2.
(ii) $T_{2}$ is "regulating" if $\alpha-(\nu / 2)>0$, i.e.

$$
\begin{aligned}
& T_{2} \geq 0 \quad \text { if } \quad\left|w_{t}(1, t)\right| \leq \sqrt{\frac{\alpha-\frac{\nu}{2}}{\beta}} \\
& T_{2}<0 \quad \text { if } \quad\left|w_{t}(1, t)\right|>\sqrt{\frac{\alpha-\frac{\nu}{2}}{\beta}}
\end{aligned}
$$

Thus energy is increasing if velocity is small, and decreasing if velocity is large.
(iii) $T_{2}$ is dissipative, i.e. $T_{2} \leq 0$, if $\alpha-(\nu / 2) \leq 0$.

To study (12), we now convert it into an equivalent first-order hyperbolic system: letting

$$
\begin{equation*}
(U, V)=\left(w_{x}, w_{t}\right) \tag{15}
\end{equation*}
$$

from $(12)_{1}$ we have

$$
\begin{align*}
\frac{\partial}{\partial t}\left[\begin{array}{l}
U \\
V
\end{array}\right] & =\left[\begin{array}{cc}
0 & 1 \\
1 & -\nu
\end{array}\right] \frac{\partial}{\partial x}\left[\begin{array}{l}
U \\
V
\end{array}\right] \\
& \equiv A \frac{\partial}{\partial x}\left[\begin{array}{l}
U \\
V
\end{array}\right] \tag{16}
\end{align*}
$$

The above hyperbolic system is not yet diagonalized. We first determine the two eigenvalues $\lambda_{1}$ and $\lambda_{2}$ and corresponding eigenvectors:

$$
\begin{align*}
& \lambda_{1}=\rho_{1}(\nu) \equiv \frac{-\nu+\sqrt{4+\nu^{2}}}{2}>0, \quad\left[\begin{array}{c}
1 \\
\rho_{1}(\nu)
\end{array}\right] \\
& \lambda_{2}=-\rho_{2}(\nu) \equiv-\frac{\nu+\sqrt{4+\nu^{2}}}{2}<0, \quad\left[\begin{array}{c}
1 \\
-\rho_{2}(\nu)
\end{array}\right] \tag{17}
\end{align*}
$$

Note that

$$
\begin{gather*}
\rho_{1}(\nu) \rho_{2}(\nu)=1, \quad \rho_{2}(\nu)-\rho_{1}(\nu)=\nu>0 \\
\rho_{1}(\nu)+\rho_{2}(\nu)=\sqrt{\nu^{2}+4} \tag{18}
\end{gather*}
$$

Let

$$
P(\nu)=\left[\begin{array}{cc}
1 & 1  \tag{19}\\
\rho_{1}(\nu) & -\rho_{2}(\nu)
\end{array}\right]
$$

Then

$$
P(\nu)^{-1}=\frac{1}{\rho_{1}(\nu)+\rho_{2}(\nu)}\left[\begin{array}{rr}
\rho_{2}(\nu) & 1  \tag{20}\\
\rho_{1}(\nu) & -1
\end{array}\right]
$$

Define

$$
\left[\begin{array}{l}
u(x, t)  \tag{21}\\
v(x, t)
\end{array}\right]=P^{-1}\left[\begin{array}{l}
U(x, t) \\
V(x, t)
\end{array}\right]
$$

Then from (16) and (19)-(21),

$$
\begin{align*}
{\left[\begin{array}{l}
u \\
v
\end{array}\right]_{t} } & =P^{-1}\left[\begin{array}{l}
U \\
V
\end{array}\right]_{t}=P^{-1}\left(A\left[\begin{array}{l}
U \\
V
\end{array}\right]_{x}\right) \\
& =P^{-1} A P\left[\begin{array}{l}
u \\
v
\end{array}\right]_{x} \tag{22}
\end{align*}
$$

$$
\left[\begin{array}{l}
u \\
v
\end{array}\right]_{t}=\left[\begin{array}{cc}
\rho_{1}(\nu) & 0 \\
0 & -\rho_{2}(\nu)
\end{array}\right]\left[\begin{array}{l}
u \\
v
\end{array}\right]_{x}
$$

and thus (16) is diagonalized to (23). Note that $\rho_{1}(\nu)$ and $\rho_{2}(\nu)$ are the speeds of two families of characteristics, and they are unequal (if $\nu \neq 0$ ).

Relations (15) and (21), written in long hand, give

$$
\begin{align*}
& u=\frac{1}{\rho_{1}(\nu)+\rho_{2}(\nu)}\left[\rho_{2}(\nu) w_{x}+w_{t}\right] \\
& v=\frac{1}{\rho_{1}(\nu)+\rho_{2}(\nu)}\left[\rho_{1}(\nu) w_{x}-w_{t}\right] \tag{23}
\end{align*}
$$

The inverse relations are

$$
\begin{align*}
& w_{x}=u+v \\
& w_{t}=\rho_{1}(\nu) u-\rho_{2}(\nu) v \tag{24}
\end{align*}
$$

From (23), we can apply the method of characteristics to get

$$
\begin{align*}
& u(x, t)=\phi\left(x+\rho_{1}(\nu) t\right) \\
& v(x, t)=\psi\left(x-\rho_{2}(\nu) t\right) \tag{25}
\end{align*}
$$

for some functions $\phi$ and $\psi$ (of a single variable). The above is valid for all $(x, t) \in \mathbb{R}^{2}$ if there were no boundary conditions imposed at $x=0$ and $x=1$. But we do have boundary conditions here, so (25) is not valid for all $(x, t) \in \mathbb{R}^{2}$ and we need to take boundary conditions into account, as given below.

The initial conditions for $u$ and $v$, from (12) ${ }_{4}$ and (23), are

$$
\begin{align*}
& u(x, 0)=u_{0}(x)=\frac{1}{\rho_{1}(\nu)+\rho_{2}(\nu)}\left[\rho_{2}(\nu) w_{0}^{\prime}(x)+w_{1}(x)\right] \\
& v(x, 0)=v_{0}(x)=\frac{1}{\rho_{1}(\nu)+\rho_{2}(\nu)}\left[\rho_{1}(\nu) w_{0}^{\prime}(x)-w_{1}(x)\right] \tag{26}
\end{align*}
$$

The boundary condition at $x=0$, i.e. $(12)_{2}$, follows from (24) ${ }_{1}$ :

$$
u(0, t)+v(0, t)=0, \quad t>0,
$$

or

$$
\begin{equation*}
v(0, t)=-u(0, t) \equiv G(u(0, t)), \quad t>0 . \tag{27}
\end{equation*}
$$

Even though $G(u)=-u$ is a simple function as defined in (27), let us stick with this seemingly more general notation $G$ so that it may also represent other maps in Sec. 7; see (69) and (71).

The boundary condition at $x=1$, i.e. $(12)_{3}$, gives

$$
\begin{equation*}
u+v=\alpha\left(\rho_{1} u-\rho_{2} v\right)-\beta\left(\rho_{1} u-\rho_{2} v\right)^{3} . \tag{28}
\end{equation*}
$$

Denote

$$
\begin{equation*}
X=\rho_{1} u-\rho_{2} v . \tag{29}
\end{equation*}
$$

Then (29) gives

$$
\begin{equation*}
u+v=\frac{1}{\rho_{1}} X+\frac{\rho_{2}}{\rho_{1}} v+v=\rho_{2} X+\left(\rho_{2}^{2}+1\right) v, \tag{30}
\end{equation*}
$$

and from (28) we obtain

$$
\begin{equation*}
\beta X^{3}+\left(\rho_{2}-\alpha\right) X+\left(\rho_{2}^{2}+1\right) v=0 . \tag{31}
\end{equation*}
$$

Under the condition

$$
\begin{equation*}
\rho_{2}(\nu)-\alpha=\frac{\sqrt{4+\nu^{2}}+\nu}{2}-\alpha \geq 0, \tag{32}
\end{equation*}
$$

it is known (see [Chen et al., 1998a, (2.1)-(2.8)]) from Cardan's formula that for each given $v \in \mathbb{R}$, there exists a unique real solution $g_{\nu}(v)$ (a singlevalued function of $v$ ) such that

$$
\begin{align*}
\beta g_{\nu}(v)^{3}+\left(\rho_{2}-\alpha\right) g_{\nu}(v)+ & \left(\rho_{2}^{2}+1\right) v=0 \\
& \left(X=g_{\nu}(v)\right) . \tag{33}
\end{align*}
$$

On the other hand, if (32) is violated, then $g_{\nu}(v)$ is multivalued: for each given real $v$, there will exist at least one, but as many as three, real values $g_{\nu}(v)$ satisfying (33).

Remark 1.4. In (33), the parameter $\beta$ plays the role of "scaling". This can be seen as follows. Express explicitly the dependence of $g_{\nu}(v)$ on $\beta$ as

$$
\begin{equation*}
g_{\nu}(v)=g_{\nu}(v, \beta) \tag{34}
\end{equation*}
$$

We try to see if the dependence on $\beta$ in (34) can somehow be "eliminated". Let us try $\beta^{\mu_{1}} g_{\nu}\left(\beta^{\mu_{2}} v\right.$, $\beta$ ) for $g_{\nu}(v)$ in (33):

$$
\begin{aligned}
\beta \cdot & {\left[\beta^{\mu_{1}} g_{\nu}\left(\beta^{\mu_{2}} v, \beta\right)\right]^{3}+\left(\rho_{2}-\alpha\right) \beta^{\mu_{1}} g_{\nu}\left(\beta^{\mu_{2}} v, \beta\right) } \\
& +\left(\rho_{2}^{2}+1\right) \beta^{\mu_{2}} v=0
\end{aligned}
$$

by setting $\mu_{1}=\mu_{2}=-1 / 2$, the above becomes

$$
\beta^{-1 / 2}\left[g_{\nu}^{3}+\left(\rho_{2}-\alpha\right) g_{\nu}+\left(\rho_{2}^{2}+1\right) v\right]=0,
$$

i.e.

$$
\beta^{-1 / 2} g_{\nu}\left(\beta^{-1 / 2} v, \beta\right)=g_{\nu}(v, 1)
$$

where $g_{\nu}(v, 1)$ is independent of $\beta$. The factor $\beta^{-\frac{1}{2}}$ will appear in many expressions (cf. Lemma 2.1, Lemmas 2.3-2.6, etc.) in Sec. 2. It is useful to recognize here first that $\beta^{-\frac{1}{2}}$ is just a scaling factor.

From (30), we therefore obtain an implicit representation of $u$ in terms of $v$ :

$$
\begin{equation*}
u=\rho_{2}\left[\rho_{2} v+g_{\nu}(v)\right]=F_{\nu}(v), \quad v \in \mathbb{R} . \tag{35}
\end{equation*}
$$

From now on, without further mention, we will always assume that $\alpha, \beta, \nu>0$ also satisfy condition (32). Since $\nu>0$, we have $\rho_{2}(\nu)>1$. From (32),

$$
\begin{equation*}
0<\alpha \leq \rho_{2} \tag{36}
\end{equation*}
$$

Therefore $\alpha$ is allowed to take values larger than one. (In contrast, the range of $\alpha$ in papers [Chen et al., 1998a-1998d] was restricted to $0<\alpha \leq 1$ in order to have the uniqueness of real solutions of cubic equations analogous to (33) therein.)

Remark 1.5. For $0<\nu<2$, we have
$\rho_{1}(\nu)=1-\frac{\nu}{2}+O\left(\nu^{2}\right), \quad \rho_{2}(\nu)=1+\frac{\nu}{2}+O\left(\nu^{2}\right)$.
For $\nu>2$, we have

$$
\rho_{1}(\nu)=\frac{1}{\nu}+O\left(\nu^{-2}\right), \quad \rho_{2}(\nu)=\nu+O(1) .
$$

Return to (23). The $u$-component remains constant on each characteristic $x+\rho_{1}(\nu) t=c=$ a positive constant, which moves leftward as $t$ increases because $\rho_{1}(\nu)>0$ by $(17)_{1}$. Each characteristic


Fig. 1. Characteristics and reflection relations for the diagonalized first-order hyperbolic system (23).
as such will intersect the left boundary $x=0$ at time $t=c / \rho_{1}(\nu)$, thereafter a reflection at $x=0$ takes place. The reflection relation is precisely (27). Thus, the incoming characteristic $x+\rho_{1}(\nu) t=c$ and the $u$-component are reflected to yield an outgoing characteristic $x-\rho_{2}(\nu) t=-\rho_{2}^{2}(\nu) c$ moving rightward with increasing $t$; on this characteristic the
$v$-component remains constant. The same can be said about reflection taking place at the right end $x=1$, where the reflection relation is given by (35). Graphically, this is illustrated in Fig. 1.

From these reflection relations, by ray tracing it is straightforward to derive the following explicit representations of $u$ and $v$ : for $t=k\left(\rho_{1}+\rho_{2}\right)+\tau$, $k=0,1,2, \ldots, 0 \leq \tau<\rho_{1}+\rho_{2}$ and $0 \leq x \leq 1$,

$$
\begin{align*}
& u(x, t)= \begin{cases}\left(F_{\nu} \circ G\right)^{k}\left(u_{0}\left(x+\rho_{1} \tau\right)\right), & \tau \leq \rho_{2}(1-x), \\
F_{\nu} \circ\left(G \circ F_{\nu}\right)^{k}\left(v_{0}\left(1+\rho_{2}^{2}-\rho_{2}^{2}\left(x+\rho_{1} \tau\right)\right)\right), & \rho_{2}(1-x)<\tau \leq \rho_{2}\left(1+\rho_{1}^{2}-x\right), \\
\left(F_{\nu} \circ G\right)^{k+1}\left(u_{0}\left(x+\rho_{1} \tau-1-\rho_{1}^{2}\right)\right), & \rho_{2}\left(1+\rho_{1}^{2}-x\right)<\tau<\rho_{1}+\rho_{2},\end{cases}  \tag{37}\\
& v(x, t)= \begin{cases}\left(G \circ F_{\nu}\right)^{k}\left(v_{0}\left(x-\rho_{2} \tau\right)\right), & \tau \leq \rho_{1} x \\
G \circ\left(F_{\nu} \circ G\right)^{k}\left(u_{0}\left(-\rho_{1}^{2}\left(x-\rho_{2} \tau\right)\right)\right), & \rho_{1} x<\tau \leq \rho_{1}\left(x+\rho_{2}^{2}\right), \\
\left(F_{\nu} \circ G\right)^{k+1}\left(v_{0}\left(x-\rho_{2} \tau+1+\rho_{2}^{2}\right)\right), & \rho_{1}\left(x+\rho_{2}^{2}\right)<\tau<\rho_{1}+\rho_{2} .\end{cases} \tag{38}
\end{align*}
$$

Note that in the above, for example, $\left(G \circ F_{\nu}\right)^{k}$ denotes the $k$-times iterative composition of $G \circ F_{\nu}$ with itself (rather than the $k$ th power of the function $G \circ F_{\nu}$ ). From these explicit representations, we see that the $u$-component and/or the $v$-component show(s) chaotic behavior if and only if the map(s) $G \circ F_{\nu}$ and/or $F_{\nu} \circ G$ is/are chaotic. We will make more precise what we mean by being chaotic or nonchaotic of the system (12) from that of $u$ and $v$ based upon certain topological conjugacy; see Sec. 5.

The organization of the paper is given as follows. In Sec. 2, we list various basic properties of the map $G \circ F_{\nu}$. In Sec. 3, we establish the route to
chaos: period-doubling, homoclinic orbits and bifurcations, and invariant Cantor sets. In Sec. 4, we prove the topological conjugacy between $(u, v)$ and $\left(w_{x}, w_{t}\right)$. In Sec. 5 , we examine the differentiability of the solution, from which chaotic behavior of $w$ itself may also be deduced. Numerical simulations of chaotic vibration are given in Sec. 6. Additional discussions are made in the final Sec. 7.

## 2. Basic Properties of the $\operatorname{Map} G \circ \boldsymbol{F}_{\nu}$

It is easy to check that $F_{\nu}$ is odd: $F_{\nu}(-v)=-F_{\nu}(v)$.


Fig. 2. The graph of $u=G F_{\nu}(v)$, with $\alpha=4 / 3, \beta=1$, $\nu=1.24$.

Also, because $G(v)=-v$, we see that $G \circ F_{\nu}(v)=$ $F_{\nu} \circ G(v)=-F_{\nu}(v)$. Therefore, in (37) and (38) investigation of the periodic and chaotic behavior of $G \circ F_{\nu}$ alone suffices here because $F_{\nu} \circ G$ is the same function. Throughout most of the discussion in this section, we will regard $\alpha, \beta$ as given fixed constants, and $\nu$ as a varying parameter.

As a visual aid, we first plot a graph of $G \circ F_{\nu}$ in Fig. 2, for $\alpha=4 / 3, \beta=1$ and $\nu=1.24$. Let us begin to analyze the properties of $G \circ F_{\nu}$, below.

Lemma 2.1 (Fixed Points). Let $\nu>0,0<\alpha \leq$ $\rho_{2}(\nu)$ and $\beta>0$. Then $G \circ F_{\nu}$ has exactly three fixed points $0, v_{0}$ and $-v_{0}$, where $v_{0}=v_{0}(\nu)=$ $\left[1 /\left(\rho_{1}+\rho_{2}\right)\right] \sqrt{\alpha / \beta}$.

Proof. If

$$
\begin{aligned}
v_{0} & =G \circ F_{\nu}\left(v_{0}\right)=-F_{\nu}\left(v_{0}\right) \\
& =-\rho_{2}^{2} v_{0}-\rho_{2} g_{\nu}\left(v_{0}\right)
\end{aligned}
$$

then

$$
\begin{equation*}
g_{\nu}\left(v_{0}\right)=-\frac{1}{\rho_{2}}\left(1+\rho_{2}^{2}\right) v_{0}=-\left(\rho_{1}+\rho_{2}\right) v_{0} \tag{39}
\end{equation*}
$$

and by (33),
$-\beta\left(\rho_{1}+\rho_{2}\right)^{3} v_{0}^{3}-\left(\rho_{1}+\rho_{2}\right)\left(\rho_{2}-\alpha\right) v_{0}+\left(\rho_{2}^{2}+1\right) v_{0}=0$,

$$
\begin{aligned}
\beta\left(\rho_{1}+\rho_{2}\right)^{3} v_{0}^{2} & =\rho_{2}^{2}+1-\left(\rho_{1}+\rho_{2}\right)\left(\rho_{2}-\alpha\right) \\
& =\alpha\left(\rho_{1}+\rho_{2}\right)>0 .
\end{aligned}
$$

Hence

$$
\begin{equation*}
v_{0}= \pm \frac{1}{\rho_{1}+\rho_{2}} \sqrt{\frac{\alpha}{\beta}} . \tag{40}
\end{equation*}
$$

Note that the graph of $u=G \circ F_{\nu}(v)$ intersects the diagonal $u-v=0$ at exactly the three points $\left(-v_{0},-v_{0}\right),(0,0)$ and $\left(v_{0}, v_{0}\right)$.

Lemma 2.2 (Derivative Formulas). Let $\nu>0$, $0<\alpha \leq \rho_{2}(\nu), \beta>0$, where $\nu$ is a varying parameter but $\alpha$ and $\beta$ are given and fixed (as long as $\left.\alpha \in\left(0, \rho_{2}(\nu)\right]\right)$. Define $f(v, \nu)=G \circ F_{\nu}(v)=$ $-\rho_{2}(\nu)\left[\rho_{2}(\nu) v+g_{\nu}(v)\right]$, where $g_{\nu}$ is defined through (33). Then

$$
\begin{equation*}
\frac{\partial}{\partial v} f(v, \nu)=-\rho_{2}^{2}+\rho_{2} \cdot \frac{\rho_{1}^{2}+1}{D} \tag{41}
\end{equation*}
$$

$$
\begin{align*}
\frac{\partial}{\partial \nu} f(v, \nu)= & -\rho_{2}^{\prime}\left[2 \rho_{2} v+g_{\nu}(v)\right] \\
& +\rho_{2} \rho_{2}^{\prime} \cdot \frac{2 \rho_{2} v+g_{\nu}(v)}{D},  \tag{42}\\
\frac{\partial^{2}}{\partial \nu \partial v} f(v, \nu)= & -\rho_{2}^{\prime}\left(2 \rho_{2}-\frac{\rho_{2}^{2}+1}{D}\right) \\
& +\rho_{2} \rho_{2}^{\prime}\left[\frac{2 \rho_{2} D-\left(\rho_{2}^{2}+1\right)}{D^{2}}\right. \\
& \left.+\frac{6 \beta\left(\rho_{2}^{2}+1\right)\left(2 \rho_{2} v+g_{\nu}(v)\right) g_{\nu}(v)}{D^{3}}\right] \tag{43}
\end{align*}
$$

$$
\begin{align*}
\frac{\partial^{2}}{\partial v^{2}} f(v, \nu) & =\frac{6 \beta \rho_{2}\left(\rho_{2}^{2}+1\right)^{2} g_{\nu}(v)}{D^{3}}  \tag{44}\\
\frac{\partial^{3}}{\partial v^{3}} f(v, \nu) & =\frac{24 \beta \rho_{2}\left(\rho_{2}^{2}+1\right)^{4}}{D^{4}} \tag{45}
\end{align*}
$$

where

$$
\begin{gathered}
D=D(v, \nu, \alpha, \beta)=3 \beta g_{\nu}^{2}(v)+\rho_{2}-\alpha, \\
\rho_{2}^{\prime}=\rho_{2}^{\prime}(\nu)=\frac{\nu+\sqrt{4+\nu^{2}}}{2 \sqrt{4+\nu^{2}}} .
\end{gathered}
$$

Proof. We will only verify (41); the rest can be done in a similar way. By the definition of $f(v, \nu)$, we have

$$
\begin{equation*}
\frac{\partial}{\partial v} f(v, \nu)=-\rho_{2}^{2}-\rho_{2} g_{\nu}^{\prime}(v) \tag{46}
\end{equation*}
$$

By differentiating (33) with respect to $v$, we get

$$
3 \beta g_{\nu}^{2}(v) g_{\nu}^{\prime}(v)+\left(\rho_{2}-\alpha\right) g_{\nu}^{\prime}(v)+\left(\rho_{2}^{2}+1\right)=0
$$

$$
\begin{equation*}
g_{\nu}^{\prime}(v)=-\frac{\rho_{2}^{2}+1}{3 \beta g_{\nu}^{2}(v)+\rho_{2}-\alpha} \tag{47}
\end{equation*}
$$

Substitution of (47) into (46) gives us (41).
Corollary 2.1. The origin is a repelling fixed point of the map $G \circ F_{\nu}$.

Proof. At $v=0$, we have $g_{\nu}(0)=0$ and $g_{\nu}^{\prime}(0)=$ $-\left(\rho_{2}^{2}+1\right) /\left(\rho_{2}-\alpha\right)<0$. By (46) and (47),

$$
\begin{aligned}
{\left[G \circ F_{\nu}\right]^{\prime}(0) } & =-\rho_{2}^{2}+\frac{\rho_{2}^{3}+\rho_{2}}{\rho_{2}-\alpha}=\frac{\rho_{2}\left(\alpha \rho_{2}+1\right)}{\rho_{2}-\alpha} \\
& =1+\frac{\alpha\left(1+\rho_{2}^{2}\right)}{\rho_{2}-\alpha}>1
\end{aligned}
$$

Lemma 2.3 (Local Maximum, Minimum and Piecewise Monotonicity). Let $\nu>0,0<\alpha \leq \rho_{2}(\nu)$, and $\beta>0$ be given. Then $G \circ F_{\nu}$ has local extremal values

$$
\begin{align*}
& m=G \circ F_{\nu}\left(-v_{c}\right)=-\frac{2}{3} \frac{1+\alpha \rho_{2}}{\rho_{1}+\rho_{2}} \sqrt{\frac{1+\alpha \rho_{2}}{3 \beta \rho_{2}}} \\
& M=G \circ F_{\nu}\left(v_{c}\right)=\frac{2}{3} \frac{1+\alpha \rho_{2}}{\rho_{1}+\rho_{2}} \sqrt{\frac{1+\alpha \rho_{2}}{3 \beta \rho_{2}}}=-m, \tag{48}
\end{align*}
$$

where

$$
v_{c}=\frac{3 \rho_{2}^{2}-2 \alpha \rho_{2}+1}{3 \rho_{2}\left(\rho_{2}^{2}+1\right)} \sqrt{\frac{1+\alpha \rho_{2}}{3 \beta \rho_{2}}} \quad \text { and }-v_{c}
$$

are critical points of $G \circ F_{\nu}$, and $m, M$ are, respectively, the local minimum and maximum of $G \circ F_{\nu}$. The function $G \circ F_{\nu}$ is strictly decreasing on $\left(-\infty,-v_{c}\right)$ and $\left(v_{c}, \infty\right)$, but strictly increasing on $\left(-v_{c}, v_{c}\right)$.

Proof. Use (41), (35), etc., and carry out the computations.

Lemma 2.4 ( $v$-Axis Intercepts). Let $\nu>0,0<$ $\alpha \leq \rho_{2}(\nu)$, and $\beta>0$ be given. Then $u=G \circ F_{\nu}(v)$ intersects the $v$-axis at abscissas $\pm v_{I}$, where $v_{I}=$ $\left(1 / \rho_{2}\right) \sqrt{\left(1+\alpha \rho_{2}\right) / \beta \rho_{2}}$.

Proof. Set $G \circ F_{\nu}\left(v_{I}\right)=0$ and solve for $v_{I}$.
Lemma 2.5 (Intersections with the Diagonal Line $u+v=0$ ). Let $\nu>0,0<\alpha \leq \rho_{2}(\nu)$ and $\beta>0$
be given. Then the graph of $u=G \circ F_{\nu}(v)$ intersects the line $u+v=0$ at exactly the following three points:

$$
(-\tilde{v}, \tilde{v}), \quad(0,0), \quad(\tilde{v},-\tilde{v})
$$

where

$$
\tilde{v}=\tilde{v}(\nu)=\frac{1}{\rho_{2}-\rho_{1}} \sqrt{\frac{1+\alpha\left(\rho_{2}-\rho_{1}\right)}{\beta\left(\rho_{2}-\rho_{1}\right)}} .
$$

Proof. Set $G \circ F_{\nu}(\tilde{v})=-\tilde{v}$ and solve for $\tilde{v}$.
Lemma 2.6 (Bounded Invariant Interval $\mathcal{I}$ ). Let $\nu>0,0<\alpha \leq \rho_{2}(\nu)$ and $\beta>0$ be given, such that

$$
\begin{align*}
M & =\frac{2}{3} \frac{1+\alpha \rho_{2}}{\rho_{1}+\rho_{2}} \sqrt{\frac{1+\alpha \rho_{2}}{3 \beta \rho_{2}}} \leq \tilde{v} \\
& =\frac{1}{\rho_{2}-\rho_{1}} \sqrt{\frac{1+\alpha\left(\rho_{2}-\rho_{1}\right)}{\beta\left(\rho_{2}-\rho_{1}\right)}} \tag{49}
\end{align*}
$$

Then the iterates of every point in the set $U \equiv$ $(-\infty, \tilde{v}) \cup(\tilde{v}, \infty)$ by $G \circ F_{\nu}$ escape to $\pm \infty$, while $\mathcal{I} \equiv \mathbb{R} \backslash \bar{U}=[-\tilde{v}, \tilde{v}]$ is an invariant interval of $G \circ F_{\nu}$.

Proof. Obvious; see Fig. 2, for example.
The set $U$ in Lemma 2.6 is the unstable set, and the set $\mathcal{I}$ is the bounded invariant set. When the condition (49) is violated, bounded invariant interval $\mathcal{I}$ no longer exists. Instead, we have a bounded Cantor-like invariant set; see Sec. 3.

## 3. Routes to Chaos

First, we show that there is a period-doubling cascade.

Theorem 3.1 (Period-Doubling Bifurcation Theorem for the $\left.\operatorname{Map} G \circ F_{\nu}\right)$. Let $\alpha, \beta>0$ be given such that $0<\alpha \leq \sqrt{2}$. Define $\bar{\nu}_{1, \alpha}$ by

$$
\bar{\nu}_{1, \alpha}=\left\{\begin{align*}
& \varepsilon, \text { if } 0<\alpha<1, \text { where } \varepsilon \text { is any small }  \tag{50}\\
& \text { positive number, } \\
& r, \text { if } \alpha \geq 1, \text { where } r \text { is the unique } \\
& \text { positive solution of } \rho_{2}(r)=\alpha
\end{align*}\right.
$$

Let $\nu \in\left[\bar{\nu}_{1, \alpha}, \infty\right)$ be a varying parameter. Define $f(v, \nu)=G F_{\nu}(v)$. Then
(i) $\alpha$ satisfies $0<\alpha \leq \rho_{2}(\nu)$ for all $\nu \in\left[\bar{\nu}_{1, \alpha}, \infty\right)$.
(ii) $v_{0}(\nu)=\left[1 /\left(\rho_{1}(\nu)+\rho_{2}(\nu)\right] \sqrt{\alpha / \beta}\right.$ is a curve of fixed points of $f: f\left(v_{0}(\nu), \nu\right)=v_{0}(\nu)$.
(iii) For $\nu_{0}=1 / \alpha$, we have $\nu_{0}>\bar{\nu}_{1, \alpha}, \rho_{2}\left(\nu_{0}\right) \geq \alpha$, and

$$
\begin{equation*}
\left.\frac{\partial}{\partial v} f(v, \nu)\right|_{\substack{v=v_{0}(\nu) \\ \nu=\nu_{0}}}=-1 \tag{51}
\end{equation*}
$$

(iv) For $\nu=\nu_{0}=1 / \alpha$ and $v=v_{0}(\nu)$, we have

$$
\begin{align*}
A & \left.\equiv\left[\frac{\partial^{2} f}{\partial \nu \partial v}+\frac{1}{2}\left(\frac{\partial f}{\partial \nu}\right) \frac{\partial^{2} f}{\partial v^{2}}\right]\right|_{\substack{v=v_{0}(\nu) \\
\nu=\nu_{0}}} \\
& \neq 0 \tag{52}
\end{align*}
$$

(v) For $\nu=\nu_{0}=1 / \alpha$ and $v=v_{0}(\nu)$, we have

$$
B=\left.\left[\frac{1}{6} \frac{\partial^{3} f}{\partial v^{3}}+\frac{1}{4}\left(\frac{\partial^{2} f}{\partial v^{2}}\right)^{2}\right]\right|_{\substack{v=v_{0}(\nu) \\ \nu=\nu_{0}}}>0
$$

Consequently, there is period-doubling bifurcation at $(v, \nu)=\left(v_{0}(\nu), 1 / \alpha\right)$. The stability type of the bifurcated period-2 orbit is attracting.

## Proof

(i) Since $\rho_{2}(\nu)=\left[\sqrt{\nu^{2}+4}+\nu\right] / 2$ is an increasing function of $\nu$ with range $[1, \infty)$, we see that if $\alpha \geq 1, \rho_{2}(\nu)=\alpha$ has a unique positive solution $\bar{\nu}_{1, \alpha}$ such that $\rho_{2}(\nu) \geq \alpha$ if $\nu \geq \bar{\nu}_{1, \alpha}$.
(ii) See Lemma 2.1.
(iii) Let us find $\nu_{0}$ satisfying (51):

$$
\left.\frac{\partial}{\partial v} f(v, \nu)\right|_{\substack{v=v_{0}(\nu) \\ \nu=\nu_{0}}}=-\rho_{2}^{2}-\rho_{2} g_{\nu_{0}}^{\prime}\left(v_{0}\left(\nu_{0}\right)\right)
$$

or

$$
\begin{align*}
g_{\nu_{0}}^{\prime}\left(v_{0}\left(\nu_{0}\right)\right) & =\frac{1}{\rho_{2}}\left(1-\rho_{2}^{2}\right)=\rho_{1}-\rho_{2} \\
& =-\frac{\rho_{2}^{2}+1}{3 \beta g_{\nu_{0}}^{2}\left(v_{0}\left(\nu_{0}\right)\right)+\rho_{2}-\alpha} \\
& =-\frac{\rho_{2}^{2}+1}{3 \beta\left[-\left(\rho_{1}+\rho_{2}\right) \cdot \frac{1}{\rho_{1}+\rho_{2}} \sqrt{\frac{\alpha}{\beta}}\right]^{2}+\rho_{2}-\alpha} \quad(\text { by }(47)) \\
& =-\frac{\rho_{2}^{2}+1}{\rho_{2}+2 \alpha} \tag{53}
\end{align*}
$$

$$
\rho_{1}-\rho_{2}=-\frac{\rho_{2}^{2}+1}{\rho_{2}+2 \alpha}=-\frac{\rho_{1}+\rho_{2}}{1+2 \alpha \rho_{1}}
$$

$$
\begin{equation*}
\nu_{0}=\frac{\sqrt{4+\nu_{0}^{2}}}{1+\alpha\left(\sqrt{4+\nu_{0}^{2}}-\nu_{0}\right)}, \quad \nu_{0}=\frac{1}{\alpha} \tag{54}
\end{equation*}
$$

(iv) We substitute (42)-(44) into the defining equation (52) for $A$. Note from (iii) above that we have

$$
\begin{gathered}
\nu_{0}=\frac{1}{\alpha}, \quad g_{\nu_{0}}\left(v_{0}\left(\nu_{0}\right)\right)=-\sqrt{\frac{\alpha}{\beta}} \\
v_{0}\left(\nu_{0}\right)=\frac{\sqrt{\frac{\alpha}{\beta}}}{\rho_{1}\left(\nu_{0}\right)+\rho_{2}\left(\nu_{0}\right)}, \quad(\text { by }(40))
\end{gathered}
$$

$$
\begin{gathered}
D=2 \alpha+\frac{\sqrt{\nu_{0}^{2}+4}+\nu_{0}}{2}=\rho_{2}\left(\nu_{0}\right)+2 \alpha, \\
\rho_{2}^{\prime}\left(\nu_{0}\right)=\frac{\rho_{2}\left(\nu_{0}\right)}{\rho_{1}\left(\nu_{0}\right)+\rho_{2}\left(\nu_{0}\right)}, \\
2 \rho_{2}\left(\nu_{0}\right) v_{0}\left(\nu_{0}\right)+g_{\nu_{0}}\left(v_{0}\left(\nu_{0}\right)\right)=\nu_{0} v_{0}\left(\nu_{0}\right) .
\end{gathered}
$$

We can substitute all of the above into (42)-(44), obtaining $A=A(\alpha)$, i.e. $A$ is independent of $\beta$. We plot the graph of $A(\alpha)$ against $\alpha: 0<\alpha \leq \sqrt{2}$ in Fig. 3 and found that $A(\alpha)<0$ for any given $\alpha>0$. (This part of the proof is computer-assisted.)
(v) This is obvious because $\partial^{3} f / \partial v^{3}>0$ from (45).

The graph of $G \circ F_{\nu}$ is unimodal for $v \geq 0$. Through the renormalization argument of [Feigenbaum, 1978] and [Collet \& Tresser, 1962], we see that once period-doubling occurs, it will repeat as a cascade, leading to chaos.

The period-doubling cascade of $G \circ F_{\nu}$ can be seen in Fig. 4 for $\alpha=1 / 2$ and $\beta=1$.

If the condition $0<\alpha \leq \sqrt{2}$ in Theorem 3.1 is violated, then there is still period doubling, except that the first period doubling occurs on period-2 orbits, making them into period- 4 orbits, such as shown in Fig. 5. If $\alpha$ keeps increasing, then the first period doubling occurs on period- $2^{n}$ orbits for some larger $n$.

Next, we show the existence of homoclinic orbits.

Theorem 3.2 (Homoclinic Orbits). Define

$$
\begin{align*}
h\left(\rho_{2}(\nu)\right) & =\frac{1}{1+\alpha \rho_{2}}\left(1+\frac{1}{\rho_{2}^{2}}\right) \\
& =\frac{1}{1+\alpha \rho_{2}(\nu)}\left(1+\frac{1}{\rho_{2}^{2}(\nu)}\right) \tag{55}
\end{align*}
$$

and let $\bar{\nu}_{2, \alpha}$ be the unique positive solution of $h\left(\rho_{2}(\nu)\right)=2 / 3 \sqrt{3}$. Assume that either
(i) $\alpha>3 \sqrt{3}-1$ and $\nu \geq \bar{\nu}_{1, \alpha}$, or
(ii) $0<\alpha \leq 3 \sqrt{3}-1$ and $\nu>\min \left(\bar{\nu}_{1, \alpha}, \bar{\nu}_{2, \alpha}\right)$, where $\bar{\nu}_{1, \alpha}$ is defined as in (50).

Then the repelling fixed point 0 of the map $G \circ F_{\nu}$ has nondegenerate homoclinic orbits. Furthermore, if $\nu=\bar{\nu}_{2, \alpha} \geq \bar{\nu}_{1, \alpha}$ then the fixed point 0 has degenerate homoclinic orbits.

Proof. Homoclinic orbits exist near the origin if $M>v_{I}$; cf. Lemmas 2.3 and 2.4 for notation. (The concept of a homoclinic orbit is very much geometric in nature. This enables an easy visual confirmation from a graph such as Fig. 2.) Thus

$$
\begin{aligned}
& M>v_{I} \quad \text { if and only if } \\
& \quad \frac{2}{3} \frac{1+\alpha \rho_{2}}{\rho_{1}+\rho_{2}} \sqrt{\frac{1+\alpha \rho_{2}}{3 \beta \rho_{2}}}>\frac{1}{\rho_{2}} \sqrt{\frac{1+\alpha \rho_{2}}{\beta \rho_{2}}},
\end{aligned}
$$

or, after simplifying,

$$
\begin{equation*}
\frac{2}{3 \sqrt{3}}>\frac{1}{1+\alpha \rho_{2}}\left(1+\frac{1}{\rho_{2}^{2}}\right) . \tag{56}
\end{equation*}
$$

The RHS of (56) is the function $h$ in (55) which is a decreasing function of $\rho_{2}$ and, hence, is also a


Fig. 3. The graph of $A=A(\alpha)$, cf. (52), as a function of $\alpha$. Note that $A(\alpha)$ is negative for any $\alpha \geq 0$.


Fig. 4. The orbit diagram of $G F_{\nu}$ for $\alpha=1 / 2, \beta=1$. Note that the first period doubling occurs at $\nu=2$, agreeing with (54) of Theorem 3.1.


Fig. 5. The orbit diagram of $G \circ F_{\nu}$ for $\alpha=1.5, \beta=1$. Note that $\alpha>\sqrt{2}$ here, and the first period doubling occurs on a period-4 orbit.
decreasing function of the variable $\nu$. On $[0, \infty)$, If $\alpha$ satisfies $h \circ \rho_{2}$ takes its minimum at $\nu=0$ :

$$
\begin{aligned}
h\left(\rho_{2}(0)\right) & =\frac{1}{1+\alpha \rho_{2}(0)}\left(1+\frac{1}{\rho_{2}^{2}(0)}\right) \\
& =\frac{2}{1+\alpha} .
\end{aligned}
$$

$$
\frac{2}{3 \sqrt{3}}>\frac{2}{1+\alpha}, \quad \text { i.e. } \quad \alpha>3 \sqrt{3}-1
$$

then (56) is satisfied for all $\nu \geq \bar{\nu}_{1, \alpha}$. There then exist nondegenerate homoclinic orbits.

On the other hand, if $\alpha$ satisfies

$$
\frac{2}{3 \sqrt{3}}<\frac{2}{1+\alpha}
$$

then there exists a unique $\bar{\nu}_{2, \alpha}>0$ such that $h\left(\rho_{2}(\nu)\right)=2 / 3 \sqrt{3}$. Therefore, if $\nu>\min \left(\bar{\nu}_{1, \alpha}\right.$, $\bar{\nu}_{2, \alpha}$ ), then nondegenerate homoclinic orbits also exist.

Degenerate homoclinic orbits (and, consequently, homoclinic bifurcations) arise when $M=$ $v_{I}$. This happens if and only if $\nu=\bar{\nu}_{2, \alpha} \geq \bar{\nu}_{1, \alpha}$.

Proposition 3.1. For $\alpha, \beta, \nu>0$, let

$$
M_{\nu} \equiv \frac{2}{3} \frac{1+\alpha \rho_{2}(\nu)}{\rho_{1}(\nu)+\rho_{2}(\nu)} \sqrt{\frac{1+\alpha \rho_{2}(\nu)}{3 \beta \rho_{2}(\nu)}}
$$

be the local maximum of $G \circ F_{\nu}$ as given in (48), and let

$$
\tilde{v}(\nu) \equiv \frac{1}{\rho_{2}(\nu) \rho_{1}(\nu)} \sqrt{\frac{1+\alpha\left[\rho_{2}(\nu)-\rho_{1}(\nu)\right]}{\beta\left[\rho_{2}(\nu)-\rho_{1}(\nu)\right]}}
$$

be the value of intersection of $G \circ F_{\nu}$ with $u+v=0$ as given in Lemma 2.5. Then for each given $\alpha, \beta>0$, there exists a unique $\tilde{\nu}(\alpha)$, dependent on $\alpha$ but independent of $\beta$, such that $\rho_{2}(\tilde{\nu}(\alpha))>\alpha$, $\tilde{\nu}(\alpha)$ is strictly decreasing with respect to $\alpha$, and

$$
M_{\nu}=\left.\tilde{v}(\nu)\right|_{\nu=\tilde{\nu}(\alpha)}
$$

Also,

$$
M_{\nu}>\tilde{v}(\nu) \quad \text { for } \nu>\tilde{\nu}(\alpha)
$$

Proof. It is easy to verify that as a function of $\nu, \tilde{v}(\nu)$ is strictly decreasing, for fixed $\alpha$ and $\beta$. However, as a function of $\nu, M_{\nu}$ is not monotonic with respect to $\nu$ in general. This makes the mathematical analysis more involved. Here, we simply provide a computer-assisted proof by numerically solving $M_{\nu}=\tilde{v}(\nu)$ to determine the dependence of $\nu$ on $\alpha$. The graph of $\nu=\tilde{\nu}(\alpha)$ is plotted in Fig. 6. The conclusion follows.

Corollary 3.1 (Homoclinic Orbits and Ensuing Chaos on the Bounded Invariant Interval $\mathcal{I}$ ). Given $\alpha, \beta>0$, let $\nu>0$ satisfy
(i) $\alpha>3 \sqrt{3}-1$ and $\nu \geq \bar{\nu}_{1, \alpha}$ or $0<\alpha \leq 3 \sqrt{3}-1$ and $\nu>\min \left(\bar{\nu}_{1, \alpha}, \bar{\nu}_{2, \alpha}\right)$, where $\bar{\nu}_{1, \alpha}$ and $\bar{\nu}_{2, \alpha}$ are given in Theorems 3.1 and 3.2;
(ii) $\nu \leq \tilde{\nu}(\alpha)$, where $\tilde{\nu}(\alpha)$ is given in Proposition 3.1.

Let $\mathcal{I}=[-\tilde{v}, \tilde{v}]$ be the bounded invariant interval corresponding to $\nu$ as given in Lemma 2.6. Then $G \circ F_{\nu}: \mathcal{I} \rightarrow \mathcal{I}$ is chaotic.

When $\nu>\tilde{\nu}(\alpha), G \circ F_{\nu}$ no longer has a bounded invariant interval. An exemplar graph of


Fig. 6. The graph of $\nu=\tilde{\nu}(\alpha)$ in Proposition 3.1.


Fig. 7. The graph of $G \circ F_{\nu}$ for $\alpha=4 / 3, \beta=1, \nu=5$.
the map $G \circ F_{\nu}$ is illustrated in Fig. 7. The two horizontal lines $u= \pm \tilde{v}(\nu)$ intersect the graph of $u=G \circ F_{\nu}(v)$ at a total of six points, as can be seen from Fig. 7. We denote the ordered abscissas of these six points by

$$
-\tilde{v}_{0} \equiv-\tilde{v}(\nu),-\tilde{v}_{1},-\tilde{v}_{2}, \tilde{v}_{2}, \tilde{v}_{1}, \tilde{v}_{0},
$$

where $\tilde{v}_{1}$ and $\tilde{v}_{2}$ are defined uniquely through

$$
\begin{aligned}
& 0<\tilde{v}_{2}<\tilde{v}_{1}<\tilde{v}_{0} \\
& =\tilde{v}(\nu)=\frac{1}{\rho_{2}(\nu)-\rho_{1}(\nu)} \sqrt{\frac{1+\alpha \cdot\left(\rho_{2}(\nu)-\rho_{1}(\nu)\right)}{\beta \cdot\left(\rho_{2}(\nu)-\rho_{1}(\nu)\right)}} \\
& G \circ F_{\nu}\left(\tilde{v}_{1}\right) \\
& =G \circ F_{\nu}\left(\tilde{v}_{2}\right) \\
& =\frac{1}{\rho_{2}(\nu)-\rho_{1}(\nu)} \sqrt{\frac{1+\alpha \cdot\left(\rho_{2}(\nu)-\rho_{1}(\nu)\right)}{\beta \cdot\left(\rho_{2}(\nu)-\rho_{1}(\nu)\right)}}
\end{aligned}
$$

We then define five intervals

$$
\begin{gathered}
I_{0}=\left[-\tilde{v}_{0},-\tilde{v}_{1}\right], \quad I_{1}=\left[-\tilde{v}_{2}, \tilde{v}_{2}\right], \quad I_{2}=-I_{0}, \\
A_{0}=\left(-\tilde{v}_{1},-\tilde{v}_{2}\right), \quad A_{1}=-A_{0} .
\end{gathered}
$$

It is easy to see that

$$
\begin{aligned}
v \in \mathcal{I}= & {\left[-\tilde{v}_{0}, \tilde{v}_{0}\right], \quad\left(G \circ F_{\nu}\right)^{n}(v) \in A_{0} \cup A_{1} } \\
& \text { for some } n \in\{0,1,2, \ldots\} \\
\Rightarrow & \lim _{k \rightarrow \infty}\left|\left(G \circ F_{\nu}\right)^{n}(v)\right|=\infty .
\end{aligned}
$$

The set

$$
\begin{equation*}
\Lambda=\bigcap_{n=0}^{\infty}\left(G \circ F_{\nu}\right)^{-n} \mathcal{I} \tag{57}
\end{equation*}
$$

is a closed bounded invariant subset of the map $G \circ F_{\nu}$. For every $v \in \Lambda$, we can assign an itinerary $s(v)$ of $v$ by

$$
\begin{gathered}
s(v)=\left(s_{0} s_{1} s_{2} \cdots s_{n} \cdots\right) \\
s_{n}=\left\{\begin{array}{l}
0 \\
1 \\
2
\end{array}\right\} \text { if }\left(G \circ F_{\nu}\right)^{n} v \in\left\{\begin{array}{l}
I_{0} \\
I_{1}, n=0,1,2, \ldots \\
I_{2}
\end{array}\right.
\end{gathered}
$$

Then $s(v)$ is a ternary number. We have $s(v) \in \sum_{3}$, where

$$
\begin{aligned}
\sum_{3}= & \left\{s=\left(s_{0} s_{1} s_{2} \cdots s_{n} \cdots\right) \mid s_{j}=0,1 \text { or } 2,\right. \\
& \text { for } j=0,1,2, \ldots\} .
\end{aligned}
$$

$\sum_{3}$ is endowed with a natural metric for ternary numbers. Then standard techniques in symbolic dynamics (see [Devaney, 1989; Robinson, 1995, pp. 33-37], e.g.) give us the following.

Theorem 3.3 (Chaos on Cantor-Like Invariant Set $\Lambda)$. For given $\alpha, \beta>0$, let $\nu>\tilde{\nu}(\alpha)$. Then $\Lambda$ defined in (57) is a Cantor set with measure zero, and the map $G \circ F_{\nu}$ on $\Lambda$ is topologically conjugate
to the shift map on $\sum_{3}$. Consequently, $G \circ F_{\nu}$ is chaotic on $\Lambda$.

## 4. Topological Conjugacy Between $(u, v)$ and $\left(w_{x}, w_{t}\right)$

In Sec. 3, we have determined the parameter ranges
in which $u$ and $v$ components display several types of chaotic behavior. In this section, we show that the gradient $\left(w_{x}, w_{t}\right)$ is topologically conjugate to $(u, v)$ and, therefore, is chaotic if and only if $(u, v)$ is. The proof given below is a generalization of the one in [Chen et al., 1998a, Sec. 5].

From the reflection of characteristics as shown in Fig. 2 and from (24), for any $t \geq 0$, we have

$$
\begin{aligned}
w_{x}\left(1, t+\left(\rho_{1}+\rho_{2}\right)\right)= & u\left(1, t+\left(\rho_{1}+\rho_{2}\right)\right)+v\left(1, t+\left(\rho_{1}+\rho_{2}\right)\right) \\
= & F_{\nu}\left(v\left(1, t+\left(\rho_{1}+\rho_{2}\right)\right)-u\left(0, t+\rho_{2}\right)\right. \\
= & F_{\nu}\left(-u\left(0, t+\rho_{2}\right)\right)-u\left(0, t+\rho_{2}\right) \\
= & F_{\nu}(-u(1, t))-u(1, t) \\
= & {\left[-\frac{\rho_{2}}{\rho_{1}+\rho_{2}} w_{x}(1, t)-\frac{1}{\rho_{1}+\rho_{2}} w_{t}(1, t)\right] } \\
& +F_{\nu}\left(-\frac{\rho_{2}}{\rho_{1}+\rho_{2}} w_{x}(1, t)-\frac{1}{\rho_{1}+\rho_{2}} w_{t}(1, t)\right) .
\end{aligned}
$$

Similarly,

$$
\begin{aligned}
w_{t}\left(1, t+\left(\rho_{1}+\rho_{2}\right)\right)= & \rho_{1} u\left(1, t+\left(\rho_{1}+\rho_{2}\right)\right)-\rho_{2} v\left(1, t+\left(\rho_{1}+\rho_{2}\right)\right) \\
= & \rho_{2}\left[\frac{\rho_{2}}{\rho_{1}+\rho_{2}} w_{x}(1, t)+\frac{1}{\rho_{1}+\rho_{2}} w_{t}(1, t)\right] \\
& +\rho_{1} F_{\nu}\left(-\frac{\rho_{2}}{\rho_{1}+\rho_{2}} w_{x}(1, t)-\frac{1}{\rho_{1}+\rho_{2}} w_{t}(1, t)\right) .
\end{aligned}
$$

Write the above as a $2 \times 2$ iterative system

$$
\left[\begin{array}{c}
\xi_{n+1}  \tag{58}\\
\eta_{n+1}
\end{array}\right]=\left[\begin{array}{c}
-\left(\frac{\rho_{2}}{\rho_{1}+\rho_{2}} \xi_{n}+\frac{1}{\rho_{1}+\rho_{2}} \eta_{n}\right)+G F_{\nu}\left(\frac{\rho_{2}}{\rho_{1}+\rho_{2}} \xi_{n}+\frac{1}{\rho_{1}+\rho_{2}} \eta_{n}\right) \\
\rho_{2}\left(\frac{\rho_{2}}{\rho_{1}+\rho_{2}} \xi_{n}+\frac{1}{\rho_{1}+\rho_{2}} \eta_{n}\right)+\rho_{1} G F_{\nu}\left(\frac{\rho_{2}}{\rho_{1}+\rho_{2}} \xi_{n}+\frac{1}{\rho_{1}+\rho_{2}} \eta_{n}\right)
\end{array}\right]
$$

where

$$
\begin{aligned}
\xi_{n} & =w_{x}\left(1, t_{0}+n\left(\rho_{1}+\rho_{2}\right)\right), \\
\eta_{n} & =w_{t}\left(1, t_{0}+n\left(\rho_{1}+\rho_{2}\right)\right), \\
n & \geq 0, \quad t_{0} \in\left[0, \rho_{1}+\rho_{2}\right)
\end{aligned}
$$

We now show that (58) is topologically conjugate to the one-dimensional iterative map $v_{n+1}=G F_{\nu}\left(v_{n}\right)$.

Define a set

$$
\begin{aligned}
\mathcal{C}= & \left\{(\xi, \eta) \in \mathbb{R}^{2} \mid \xi=-c+G F_{\nu}(c),\right. \\
& \left.\eta=\rho_{2} c+\rho_{1} G F_{\nu}(c), \forall c \in \mathbb{R}\right\}
\end{aligned}
$$

Then $\mathcal{C}$ is a curve in $\mathbb{R}^{2}$ which is, of course, one dimensional.

Next, define a map

$$
\mathcal{F}: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}, \quad \mathcal{F}\left[\begin{array}{c}
\xi \\
\eta
\end{array}\right]=\left[\begin{array}{c}
-\left(\frac{\rho_{2}}{\rho_{1}+\rho_{2}} \xi+\frac{1}{\rho_{1}+\rho_{2}} \eta\right)+G F_{\nu}\left(\frac{\rho_{2}}{\rho_{1}+\rho_{2}} \xi+\frac{1}{\rho_{1}+\rho_{2}} \eta\right) \\
\rho_{2}\left(\frac{\rho_{2}}{\rho_{1}+\rho_{2}} \xi+\frac{1}{\rho_{1}+\rho_{2}} \eta\right)+\rho_{1} G F_{\nu}\left(\frac{\rho_{2}}{\rho_{1}+\rho_{2}} \xi+\frac{1}{\rho_{1}+\rho_{2}} \eta\right)
\end{array}\right]
$$

Then the iteration in (58) can be written as

$$
\left(\xi_{n+1}, \eta_{n+1}\right)=\mathcal{F}\left(\xi_{n}, \eta_{n}\right), \quad n=0,1,2, \ldots
$$

It is trivial to see that $\mathcal{F}: \mathbb{R}^{2} \rightarrow \mathcal{C}$ and, therefore $\mathcal{F}$ has a restriction

$$
\begin{equation*}
\left.\tilde{\mathcal{F}} \equiv \mathcal{F}\right|_{\mathcal{C}}: \mathcal{C} \rightarrow \mathcal{C} \tag{59}
\end{equation*}
$$

Let $\mathcal{G}$ be the graph of $G F_{\nu}$ :

$$
\mathcal{G} \equiv\left\{\left(v, G F_{\nu}(v)\right) \in \mathbb{R}^{2} \mid v \in \mathbb{R}\right\}
$$

Define

$$
H: \mathcal{G} \rightarrow \mathcal{G}, \quad H\left[\begin{array}{c}
v \\
G F_{\nu}(v)
\end{array}\right]=\left[\begin{array}{c}
G F_{\nu}(v) \\
\left(G F_{\nu}\right)^{2}(v)
\end{array}\right],
$$

i.e. $H$ is the diagonal operator

$$
H=\left[\begin{array}{cc}
G F_{\nu}(\cdot) & 0 \\
0 & G F_{\nu}(\cdot)
\end{array}\right] .
$$

Let $Q: \mathbb{R}^{2} \rightarrow \mathbb{R}^{2}$ be given by

$$
Q\left[\begin{array}{l}
\xi \\
\eta
\end{array}\right]=\left[\begin{array}{c}
-\xi+\eta \\
\rho_{2} \xi+\rho_{1} \eta
\end{array}\right]
$$

Then $Q$ is invertible and

$$
\left.\tilde{Q} \equiv Q\right|_{\mathcal{G}}: \mathcal{G} \rightarrow \mathcal{C} \quad \text { homeomorphically. }
$$

Using the above, we easily verify that we have the commutative diagram

$$
\begin{array}{rlll}
\tilde{\mathcal{G}} & \xrightarrow{H} \mathcal{G}  \tag{60}\\
\tilde{Q} \downarrow \\
\underset{\mathcal{C}}{\downarrow} \xrightarrow{\tilde{\mathcal{F}}} \downarrow & \downarrow \\
\mathcal{C}
\end{array} \quad \tilde{\mathcal{F}}=\tilde{Q} H \tilde{Q}^{-1}
$$

Therefore $\tilde{\mathcal{F}}$ is topologically conjugate to a diagonal map $H$ both of whose diagonal entries are the same map $G F_{\nu}$, which is chaotic for $\nu$ and $\alpha, \beta$ lying in a certain parameter range. It is in the sense of topological conjugacy (59) that the gradient ( $w_{x}(1, t)$, $\left.w_{t}(1, t)\right)$ map (59), a Poincaré section of $\left(w_{x}(x, t)\right.$, $\left.w_{t}(x, t)\right)$ at $x=1$, is chaotic.

Note that in general the state variable $w(x, t)$ does not display chaotic behavior. It only has a fractal outlook. In order to have chaotic profile of $w(x, t)$ itself, the boundary conditions must be differentiated. See the discussion in the next section.

## 5. Differentiability of Solutions

The smoothness of solutions to linear or nonlinear PDEs is always a major issue. Here, let us examine the questions of continuity and differentiability of the solutions $u, v$ and consequently, $w$. We will follow [Chen et al., 1998b, Sec. 6].

First, note from the explicit representations in (37) and (38) that $u$ and $v$ are obtained by patching up different expressions along the following characteristics:
$\left\{\begin{array}{l}\rho_{2} x+t=k\left(\rho_{1}+\rho_{2}\right), \\ \rho_{2} x+t=k\left(\rho_{1}+\rho_{2}\right)+\rho_{2}, \\ \rho_{1} x-t=-k\left(\rho_{1}+\rho_{2}\right), \\ \rho_{1} x-t=-k\left(\rho_{1}+\rho_{2}\right)-\rho_{2}, \\ k=0,1,2, \ldots, n, \ldots ; \quad x \in[0,1], \quad t \geq 0 .\end{array}\right.$
Only along these characteristics can discontinuities propagate [Courant \& Hilbert, 1962, Sec. V.1]. Therefore, we need to study how smooth the patching in (37) and (38) is across the characteristic line segments in (61).

Theorem 5.1. Let $(u, v)$ be the solution of the system (23), (26)-(28) in the sense of the method of characteristics as represented by (37) and (38). Assume that the initial conditions $u_{0}, v_{0} \in C^{m}([0,1])$ for some $m \in\{0,1,2, \ldots\}$. In addition, assume that at the left end $x=0$, we have

$$
\begin{equation*}
v_{0}^{(j)}(0)=(-1)^{j+1} \rho_{1}^{2 j} u_{0}^{(j)}(0), \quad j=0,1, \ldots, m . \tag{62}
\end{equation*}
$$

Also at the right end $x=1$, assume that we have

$$
\begin{align*}
u_{0}^{(j)}(1)= & (-1)^{j} \rho_{2}^{2 j} \cdot \mathcal{F}_{j}\left(F_{\nu}, F_{\nu}^{\prime}, \ldots, F_{\nu}^{(j)}, v_{0}(1),\right. \\
& \left.v_{0}^{\prime}(1), \ldots, v_{0}^{(j)}(1)\right), \quad j=0,1,2, \ldots, m, \tag{63}
\end{align*}
$$

where

$$
\begin{aligned}
\mathcal{F}_{0} & =F_{\nu}\left(v_{0}(1)\right), \quad \mathcal{F}_{1}=F_{\nu}^{\prime}\left(v_{0}(1)\right) v_{0}^{\prime}(1) \\
\mathcal{F}_{2}= & F_{\nu}^{\prime \prime}\left(v_{0}(1)\right) v_{0}^{\prime}(1)^{2}+F_{\nu}^{\prime}\left(v_{0}(1)\right) v_{0}^{\prime \prime}(1) \\
\mathcal{F}_{3}= & F_{\nu}^{\prime \prime \prime}\left(v_{0}(1)\right)\left[v_{0}^{\prime}(1)\right]^{3}+3 F_{\nu}^{\prime \prime}\left(v_{0}(1)\right) v_{0}^{\prime \prime}(1) v_{0}^{\prime}(1) \\
& +F_{\nu}^{\prime}\left(v_{0}(1)\right) v_{0}^{\prime \prime \prime}(1) \\
& \vdots \\
\mathcal{F}_{m}= & F_{\nu}^{(m)}\left(v_{0}(1)\right)\left[v_{0}^{\prime}(1)\right]^{m}+\cdots+F_{\nu}^{\prime}\left(v_{0}(1)\right) v_{0}^{(m)}(1) .
\end{aligned}
$$

Then the solution $(u, v)$ is $C^{m}$-continuous on the space-time domain $\{(x, t) \mid 0 \leq x \leq 1,0 \leq t \leq T\}$ for any $T>0$.

Proof. Choose any unit vector $\mathbf{a}=\left(a_{1}, a_{2}\right)$ on the $(x, t)$-plane, and let $D_{\mathrm{a}}$ be the directional derivative along a. We need to show that $\left(D_{\mathbf{a}}\right)^{j} u$ and $\left(D_{\mathbf{a}}\right)^{j} v$ are continuous across the characteristics
line segments in (61), for any $k=0,1,2, \ldots$, for $j=0,1,2, \ldots, m$.

Take, for example, $(38)_{1}$ and $(38)_{2}$ across the characteristic segment $t-\rho_{1} x=k\left(\rho_{1}+\rho_{2}\right)$ :

$$
v(x, t)=\left\{\begin{array}{l}
\left(G \circ F_{\nu}\right)^{k}\left(v_{0}\left(x-\rho_{2} \tau\right)\right)=\left(G \circ F_{\nu}\right)^{k}\left(v_{0}\left(x-\rho_{2}\left(t-k\left(\rho_{1}+\rho_{2}\right)\right)\right)\right), \\
\quad t-k\left(\rho_{1}+\rho_{2}\right) \leq \rho_{1} x \leq \rho_{1}, \\
G \circ\left(F_{\nu} \circ G\right)^{k}\left(u_{0}\left(-\rho_{1}^{2}\left(x-\rho_{2} \tau\right)\right)\right)=\left(G \circ F_{\nu}\right)^{k}\left(-u_{0}\left(-\rho_{1}^{2} x+\rho_{1}\left(t-k\left(\rho_{1}+\rho_{2}\right)\right)\right)\right), \\
\rho_{1} x<t-k\left(\rho_{1}+\rho_{2}\right) \leq \rho_{1} x+\rho_{2} .
\end{array}\right.
$$

In order to have $v(x, t) C^{m}$-continuous across $t-\rho_{1} x=k\left(\rho_{1}+\rho_{2}\right)$, we must have

$$
\begin{aligned}
D_{\mathbf{a}}^{j}\left[\left(G \circ F_{\nu}\right)^{k}\left(v_{0}\left(x-\rho_{2} t+k \rho_{2}\left(\rho_{1}+\rho_{2}\right)\right)\right)\right]= & D_{\mathbf{a}}^{j}\left[\left(G \circ F_{\nu}\right)^{k}\left(-u_{0}\left(-\rho_{1}^{2} x+\rho_{1} t-k \rho_{1}\left(\rho_{1}+\rho_{2}\right)\right)\right)\right] \\
& \text { on } t-\rho_{1} x=k\left(\rho_{1}+\rho_{2}\right), \text { for } k=0,1,2, \ldots
\end{aligned}
$$

This leads to (62).
Similarly, we obtain (63) by the $j$-times differentiation $D_{\mathbf{a}}^{j}$ of $(37)_{1}$ and $(37)_{2}$ along the characteristic segments $t+\rho_{2} x=\rho_{2}+k\left(\rho_{1}+\rho_{2}\right)$.

All the other cases in (37) and (38) lead to the same compatibility conditions (62) and (63).

Corollary 5.1. Let $w$ be the solution of (12) such that the initial conditions $\left(w_{0}, w_{1}\right)$ satisfy $w_{0} \in$
$C^{m+1}([0,1])$ and $w_{1} \in C^{m}([0,1])$ for some nonnegative integer $m$. Let $\left(u_{0}, v_{0}\right)$ be defined as in (26) and assume that ( $u_{0}, v_{0}$ ) satisfy (37) and (38). Then $w$ is $C^{m+1}$-continuous on $[0,1] \times[0, T]$ for any $T>0$.

Now, assume that the solution $w$ of (12) is $C^{1}$ continuous. Take $\partial / \partial t$ of $(12)_{1}-(12)_{3}$. We see that $W(x, t) \equiv w_{t}(x, t)$ satisfies the following system

$$
\begin{cases}W_{x x}(x, t)-\nu W_{x t}(x, t)-W_{t t}(x, t)=0, & 0<x<1, \quad t>0  \tag{64}\\ W_{x}(0, t)=0, & t>0, \\ W_{x}(1, t)=\left[\alpha-3 \beta W^{2}(1, t)\right] W_{t}(1, t), & t>0, \\ W(x, 0)=w_{1}(x), W_{t}(x, 0)=w_{0}^{\prime \prime}(x)-\nu w_{1}^{\prime}(x), & 0<x<1\end{cases}
$$

Note that the right end boundary condition $(64)_{3}$ is also a self-excitation boundary condition, analogous to the van der Pol ODE

$$
\ddot{x}+\left(\alpha-3 \beta x^{2}\right) \dot{x}+\omega_{0}^{2} x=0 .
$$

The boundary condition $(64)_{3}$ involves $W$ itself and thus causes significant complexity if the method of characteristics were to be applied to (64). Here, we bypass such a technical difficulty by regarding (64) as a differentiated solution of (12). For (64), chaotic behavior of the state variable or displacement $W(x, t)$ itself can be observed (while for (12) only the chaotic behavior of the gradient $\left(w_{x}, w_{t}\right)$ may happen).

Theorem 5.2. Let $W$ be a $C^{m}$-continuous solution of (64) for some $m \geq 0$. Then $W$ is unique. Assume that $\alpha, \beta, \nu>0$ and $\alpha \leq \rho_{2}(\nu)$ such that the
map $G \circ F_{\nu}$ is chaotic. Then $W$ displays chaotic behavior.

Proof. The uniqueness of $W$ follows from the fact that

$$
w(x, t) \equiv \int_{0}^{t} W(x, \tau) d \tau+f_{0}(x), \quad 0 \leq x \leq 1
$$

where $f_{0}(x)$ is a $C^{m+1}$-continuous function satisfying $f_{0}^{\prime \prime}=w_{0}^{\prime \prime}$ for $w_{0}^{\prime \prime}$ given in $(64)_{4}$ and appropriate compatibility conditions, is the unique solution of (12).

The reason that $W(x, t)$ can display chaotic behavior is that $W(x, t)=w_{t}(x, t)$ for $w$ satisfying (12), and by Sec. 4, $w_{t}$ displays chaotic behavior if the map $G \circ F_{\nu}$ is chaotic.

## 6. Numerical Examples

We offer numerical examples and graphics to manifest the difference in profiles between spatio-temporal-isotropic chaotic vibration in [Chen et al.,

1998b] and the nonisotropic chaotic vibration being studied here.

Example 6.1 (Isotropic Chaotic Vibration). Consider

$$
\begin{cases}w_{x x}(x, t)-w_{t t}(x, t)=0, & 0<x<1, \quad t>0  \tag{65}\\ w_{x}(0, t)=-\eta w_{t}(0, t), & t>0, \quad \eta>0 \text { is fixed } \\ w_{x}(1, t)=\alpha w_{t}(1, t)-\beta w_{t}^{3}(1, t), & t>0, \quad 0<\alpha \leq 1, \quad \beta>0 \\ w(x, 0)=w_{0}(x), w_{t}(x, 0)=w_{1}(x), & 0<x<1\end{cases}
$$

Define

$$
\begin{aligned}
& u(x, t)=\frac{1}{2}\left[w_{x}(x, t)+w_{t}(x, t)\right], \\
& v(x, t)=\frac{1}{2}\left[w_{x}(x, t)-w_{t}(x, t)\right] .
\end{aligned}
$$

Then

$$
\begin{aligned}
& u(x, 0)=\frac{1}{2}\left[w_{0}^{\prime}(x)+w_{1}(x)\right] \equiv u_{0}(x), \\
& v(x, 0)=\frac{1}{2}\left[w_{0}^{\prime}(x)-w_{t}(x)\right] \equiv v_{0}(x) .
\end{aligned}
$$

Define the map

$$
G(v) \equiv G_{\eta}(v)=-\frac{1+\eta}{1-\eta} v,
$$

then $u(x, t)$ and $v(x, t)$ satisfy (37) and (38), with $\nu=0, \rho_{1}(0)=\rho_{2}(0)=1, G \equiv G_{\eta}$ and $F_{\nu} \equiv$ $\left.F_{\nu}\right|_{\nu=0}=F_{0}$. According to [Chen et al., 1998b, Example 4.3], the map $G_{\eta} \circ F_{0}$ is chaotic when $\alpha=0.5$, $\beta=1, \eta=1.520$, for example. Therefore both $u$ and $v$ display chaotic behavior. By choosing a $C^{2}$ spline

$$
\begin{align*}
& u_{0}(x)=\frac{1}{12} \begin{cases}\frac{\left(x-x_{1}\right)^{3}}{h^{3}}, & x_{1} \leq x \leq x_{2}, \\
1+\frac{3\left(x-x_{2}\right)}{h}+\frac{3\left(x-x_{2}\right)^{2}}{h^{2}}-\frac{3\left(x-x_{2}\right)^{3}}{h^{3}}, & x_{2} \leq x \leq x_{3}, \\
1-\frac{3\left(x-x_{4}\right)}{h}+\frac{3\left(x-x_{4}\right)^{3}}{h^{2}}+\frac{3\left(x-x_{4}\right)^{3}}{h^{3}}, & x_{3} \leq x \leq x_{4}, \\
\frac{\left(x_{5}-x\right)^{3}}{h^{3}}, & x_{4} \leq x \leq x_{5} \\
0, & \text { elsewhere } \\
h=\frac{1}{6}, x_{j}=\frac{j}{6}, \quad j=1,2,3,4,5, & \\
v_{0}(x) \equiv 0,\end{cases} \tag{66}
\end{align*}
$$

we have $u, v \in C^{2}([0,1] \times[0, T])$ for any $T>0$.
The solution profiles of $u$ and $v$ are plotted in Figs. 8 and 9, respectively, for $t: 50 \cdot 2 \leq t \leq 51 \cdot 2$. (Note that here, each cycle of vibration which is the time required for a wave to travel from one end point and return, is two time units.) Even though there is chaotic vibration, there is clearly visible orderly patterns of wave propagation on the wave fronts $x+t=$ const. and $x-t=$ const. for $u$ and $v$, respectively.

Example 6.2 (Nonisotropic Chaotic Vibration). Consider now (12) instead, with $\alpha=0.5, \beta=1$ and $\nu=3.33$. Then

$$
\rho_{1}(\nu)=0.277, \quad \rho_{2}(\nu)=3.61 .
$$

Each cycle of vibration takes $\rho_{1}+\rho_{2} \approx 3.88$ time units. According to Theorem 3.2, for these parameter values, the map $G \circ F_{\nu}$ is chaotic.

Choosing the same initial conditions for $u_{0}$ and $v_{0}$ as in (66) and (67), we note that Theorem 5.1


Fig. 8. The spatiotemporal-isotropic chaotic profile of $u(x, t), 0 \leq x \leq 1,50 \cdot 2 \leq t \leq 51 \cdot 2$, for Example 6.1. Note that chaotic wave fronts propagate with $x+t=$ constant.


Fig. 9. The spatiotemporal-isotropic chaotic profile of $v(x, t), 0 \leq x \leq 1,50 \cdot 2 \leq t \leq 51 \cdot 2$, for Example 6.1. Note that chaotic wave fronts propagate with $x-t=$ constant.
is applicable, and we again have $u, v \in C^{2}([0,1] \times$ $[0, T]$ ), for any $T>0$.

The solution profiles of $u$ and $v$ are plotted in Figs. $10-15$, for $t: 0 \leq t \leq 3.88,3.88 \leq t \leq 2 \cdot 3.88$, and $50 \cdot(3.88) \leq t \leq 51 \cdot(3.88)$. The reader may
observe sharp "randomness" in every direction of space and time. Thus, nonisotropic chaotic vibration shows strong mixing of waves.

The profiles of the gradient, $w_{t}$ and $w_{x}$, are plotted in Figs. 16 and 17, respectively, for


Fig. 10. The initial spatiotemporal profile of $u(x, t), 0 \leq x \leq 1,0 \leq t \leq 3.88$, for Example 6.2.


Fig. 11. The initial spatiotemporal profile of $v(x, t), 0 \leq x \leq 1,0 \leq t \leq 3.88$, for Example 6.2.
$50 \cdot(3.88) \leq t \leq 51 \cdot(3.88)$, according to (24). The profile of $w$ itself is given in Fig. 14, with initial condition

$$
w_{0}(x)=\int_{0}^{x} u_{0}(\xi) d \xi
$$

$0 \leq x \leq 1 ;$ see (66) for $u_{0}$.

The reader may note the fractal, but nonchaotic, pattern of $w$.

## 7. Comments on Other Types of Boundary Conditions at $\boldsymbol{x}=\mathbf{0}$

Heretofore, the boundary condition at $x=0$ was


Fig. 12. The spatiotemporal profile of $u(x, t)$ during the second time-cycle, $0 \leq x \leq 1,3.88 \leq t \leq 2 \cdot(3.88)$, for Example 6.2.


Fig. 13. The spatiotemporal profile of $v(x, t)$ during the second time-cycle, $0 \leq x \leq 1,3.88 \leq t \leq 2 \cdot(3.88)$, for Example 6.2.
assumed to be homogeneous Neumann as in (12) $)_{2}$. But some other types of boundary conditions can be treated as well.

First, let us consider the Dirichlet boundary condition

$$
\begin{equation*}
w(0, t)=0, \quad t>0, \tag{68}
\end{equation*}
$$

in lieu of (12) ${ }_{2}$. The above gives

$$
\begin{aligned}
w_{t}(0, t) & =0, \\
\rho_{1}(\nu) u(0, t)-\rho_{2}(\nu) v(0, t) & =0, \quad\left(\text { by }(24)_{2}\right)
\end{aligned}
$$



Fig. 14. The spatiotemporal-nonisotropic chaotic profile of $u(x, t), 0 \leq x \leq 1,50 \cdot(3.88) \leq t \leq 51 \cdot(3.88)$, for Example 6.2. Note the sharp 2D randomness in the chaotic zone.


Fig. 15. The spatiotemporal-nonisotropic chaotic profile of $v(x, t), 0 \leq x \leq 1,50 \cdot(3.88) \leq t \leq 51 \cdot(3.88)$, for Example 6.2. Note the sharp 2D randomness in the chaotic zone.
and so the reflection relation at $x=0$ is

$$
\begin{equation*}
v(0, t)=\frac{\rho_{1}(\nu)}{\rho_{2}(\nu)} u(0, t) \equiv G_{\nu}(u(0, t)) . \tag{69}
\end{equation*}
$$

The explicit representations (37) and (38) are again applicable, wherein $G_{\nu}$ in (69) is now replacing $G$
in (27). Therefore, whether (the new) $u$ and $v$ display chaotic behavior depends on the orderly or chaotic property of the map $G_{\nu} \circ F_{\nu}$ and/or $F_{\nu} \circ G_{\nu}$. But $F_{\nu} \circ G_{\nu}$ is topologically conjugate to $F_{\nu} \circ G_{\nu}$ through the topological conjugacy

$$
\left(G_{\nu} \circ F_{\nu}\right) \circ G_{\nu}=G_{\nu} \circ\left(F_{\nu} \circ G_{\nu}\right),
$$



Fig. 16. The profile of $w_{t}(x, t), 0 \leq x \leq 1,50 \cdot(3.88) \leq t \leq 51 \cdot(3.88)$, for Example 6.2. Again note the spatiotemporalnonisotropic chaotic feature as well as the mixture of very chaotic, mildly chaotic, and nonchaotic zones here and in Fig. 17 next.


Fig. 17. The profile of $w_{x}(x, t), 0 \leq x \leq 1,50 \cdot(3.88) \leq t \leq 51 \cdot(3.88)$, for Example 6.2.
therefore investigation of $G_{\nu} \circ F_{\nu}$ alone suffices. Instead of repeating all the analysis work in Secs. 2-5, we simply point out the consequence: $G_{\nu} \circ F_{\nu}$ is not chaotic. The reason is that the map $G_{\nu}$ in (69) is very strongly contracting because by Remark 1.5,
$\rho_{1}(\nu) / \rho_{2}(\nu)$ is considerably smaller than 1 in general. Therefore, we see that the postivity of $T_{1}$ in Remark 1.3(i) due to the Neumann boundary condition is very important. For the Dirichlet condition (68), $T_{1} \equiv 0$ in (14); there is much less energy


Fig. 18. The profile of the state variable $w(x, t), 0 \leq x \leq 1,50 \cdot(3.88) \leq t \leq 51 \cdot(3.88)$, for Example 6.2. Observe the fractal, but nonchaotic, appearance of $w$.
imbalance, which thus actually fails to produce chaos.

Second, if the boundary condition at $x=0$ is energy-injecting:

$$
\begin{equation*}
w_{x}(0, t)=-\eta w_{t}(0, t), \quad \eta>0, \eta \neq \rho_{1}(\nu), \tag{70}
\end{equation*}
$$

in lieu of $(12)_{2}$, then the new reflection relation at $x=0$ is

$$
\begin{align*}
v(0, t) & =-\frac{1+\rho_{1}(\nu) \eta}{1-\rho_{2}(\nu) \eta} u(0, t) \\
& \equiv G_{\nu, \eta}(u(0, t)) \tag{71}
\end{align*}
$$

One can use this $G_{\nu, \eta}$ instead of $G$ in (27) to analyze the behavior of $G_{\nu, \eta} \circ F_{\nu}$. Again, as the parameter(s) $\eta$ and/or $\nu$ vary, we observe chaos. However, (70) may actually be regarded as a "special case" as our treatment in Secs. 2-5, where we have shown that chaos occurs with $\eta=0$ in (70), cf. (12) $)_{2}$, i.e. we do not need extra energy injection like (70) at $x=0$.

If the boundary condition at $x=0$ is homogeneous Robin:

$$
w(0, t)+\eta w_{x}(0, t)=0, \quad \eta>0, \quad \text { for all } t>0,
$$

then the treatment becomes much more difficult. No analytical results are available so far.

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