# Random Attractor of Reaction-Diffusion Hopfield Neural Networks Driven by Wiener Processes 

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#### Abstract

This paper studies the global existence and uniqueness of the mild solution for reaction-diffusion Hopfield neural networks (RDHNNs) driven by Wiener processes by applying a Schauder fixed point theorem and a priori estimate; then the random attractor for this system is also studied by constructing proper random dynamical system.


## 1. Introduction

It is well known that the dynamics of Hopfield neural networks have been deeply investigated because they have been successfully employed in many areas such as pattern recognition, associate memory, and combinatorial optimization. The diffusion effect cannot be avoided in the neural networks when electrons are moving in asymmetric electromagnetic field. So, the dynamical behavior of reactiondiffusion Hopfield neural networks (RDHNNs, for short) has been receiving much attention, recently [1-11].

But in a more realistic model, in order to describe the propagation of an electric potential in a neuron, it is sensible to include some noise in the system. In fact, a neural network can be stabilized or destabilized by certain stochastic inputs. Many scholars have been devoted to the stochastic RDHNNs such as $[2,12,13]$. Giving a deep insight into these literatures, we will find that most of these literatures consider the RDHNNs with finite dimensional Wiener processes; there are few results on the RDHNNs driven by infinite dimensional Wiener processes. However, since the neurons can be regarded as long thin cylinders, which act like electrical cables, the infinite dimensional Wiener processes are more favorable than standard Brownian motion.

On the other hand, attractor plays an important role in the long time behavior for dynamical systems. The random
attractor extends the concept of a strange attractor from deterministic to stochastic system. There has been great interest in random attractors for stochastic partial differential equations in recent decades. The random attractors are compact invariant random sets attracting all the orbits of attraction basin. They provide crucial geometric information about their asymptotic regime as $t \rightarrow \infty$. They can help us understand the chaotic behavior of the stochastic DRDHNNs and reduce the complexity, as well as providing the statistical properties of this system. When the global existence and uniqueness of the solution can be assured, many scholars pay much attention to the global stability, boundedness, and even synchronization of the RDHNNs [1-3, 11]. However, to the best of our knowledge, there is no result on the attractor for RDHNNs, let alone random attractor for stochastic RDHNNs. We hope this work can lay a solid foundation for the future research.

So, this paper is devoted to the random attractor for the following RDHNNs driven by Wiener processes:

$$
\begin{aligned}
& \mathrm{d} u_{i}(t, \mathbf{x})=\left(\sum _ { j = 1 } ^ { l } \frac { \partial } { \partial x _ { j } } \left(G_{i j}\left(\mathbf{x} \frac{\partial u_{i}}{\partial x_{j}}\right)-b_{i} u_{i}\right.\right. \\
& \left.\quad+\sum_{j=1}^{n} c_{i j} f_{j}\left(u_{j}\right)+I_{\mathrm{i}}\right) \mathrm{d} t+\sigma_{i} \mathrm{~d} W_{i}(t, \mathbf{x}),
\end{aligned}
$$

$$
\begin{align*}
& u_{i}(t, \mathbf{x})=0, \quad t \geq t_{0}, \quad \mathbf{x} \in \partial \mathcal{O} \\
& u_{i}\left(t_{0}, \mathbf{x}\right)=\phi_{i}(\mathbf{x}), \quad \mathbf{x} \in \mathcal{O} \in \mathbb{R}^{l}, i=1,2, \ldots, n \tag{1}
\end{align*}
$$

where $u_{i}(t, \mathbf{x})$ denote the potential of the cell $i$ at $t$ and $\mathbf{x} \in \mathbb{R}^{l}$. $b_{i}$ are positive constants and denote the rate with which the $i$ th unit will reset its potential to the resting state in isolation when it is disconnected from the network and external inputs at $t . c_{i j}$ are connection weights of the neural network. $f_{i}$ are the active functions of the neural network, which are continuous. $\sigma_{i}$ are the intensity of the noise. $I_{i}$ denote the $i$ th component of an external input source introduced from outside the network to the $i$ th neuron, which are constant numbers. $\mathcal{O}$ denotes an open bounded and connected subset of $\mathbb{R}^{l}$ with a sufficiently regular boundary $\partial \mathcal{O} . \nabla$ is the gradient. Initial data $\phi_{i}$ are $\mathscr{F}_{0}$-measurable and belong to $L^{2}(\mathcal{O})$, a.e. $\omega \in \Omega$.

For convenience, we rewrite system (1) in the vector form

$$
\begin{align*}
\mathrm{d} \mathbf{u}= & (\nabla \cdot(G(\mathbf{x}) \circ \nabla \mathbf{u})-\mathbf{B} \mathbf{u}+\mathbf{C} \tilde{f}(\mathbf{u})+\mathbf{I}) \mathrm{d} t \\
& +\sigma \mathrm{d} \mathbf{W} \\
\mathbf{u}(t, \mathbf{x})= & 0, \quad t \geq t_{0}, \mathbf{x} \in \partial \mathcal{O}  \tag{2}\\
\mathbf{u}\left(t_{0}, \mathbf{x}\right)= & \phi(\mathbf{x})
\end{align*}
$$

where $\mathbf{C}=\left(c_{i j}\right)_{n \times n}, \sigma=\operatorname{diag}\left\{\sigma_{1}, \sigma_{2}, \ldots, \sigma_{n}\right\} \mathbf{I}=\left(I_{1}, I_{2}\right.$, $\left.\ldots, I_{n}\right)^{T}, \mathbf{u}=\left(u_{1}, u_{2}, \ldots, u_{n}\right)^{T}, \nabla \mathbf{u}=\left(\nabla u_{1}, \nabla u_{2}, \ldots, \nabla u_{n}\right)^{T}$, $\mathbf{W}=\left(W_{1}, W_{2}, \ldots, W_{n}\right)^{T}, \widetilde{\mathbf{f}}(\mathbf{u})=\left(f_{1}\left(u_{1}\right), f_{2}\left(u_{2}\right), \ldots, f_{n}\left(u_{n}\right)\right)^{T}$ is the diagonal map, $\nabla$ is the gradient operator, $\mathbf{B}=$ $\operatorname{diag}\left(b_{1}, b_{2}, \ldots, b_{n}\right), \boldsymbol{\phi}(\mathbf{x})=\left(\phi_{1}(\mathbf{x}), \phi_{2}(\mathbf{x}), \ldots, \phi_{n}(\mathbf{x})\right)^{T}, \mathbf{G}(\mathbf{x})=$ $\left(G_{i j}\right)_{n \times l}, \mathbf{G} \circ \nabla \mathbf{u}=\left(G_{i j}\left(\partial u_{i} / \partial x_{j}\right)\right)_{n \times l}$ is the Hadamard product between matrix $\mathbf{G}$ and $\nabla \mathbf{u}$, and $\nabla \cdot$ is the divergence operator.

We will also use the following notations in the paper.
(i) $(\Omega, \mathscr{F}, \mathbb{P})$ is a complete probability space with filtration $\left\{\mathscr{F}_{t}\right\}_{t \geq 0}$ satisfying the usual conditions (see [1, 2, 14-18]).
(ii) $\mathbf{W}(t, \mathbf{x})=\mathbf{W}(t, \mathbf{x})(\omega)$ is a space-time Wiener process with values in the separable Hilbert space $K$ with $E \mathbf{W}=\mathbf{0}$ and $E(\mathbf{W}, \mathbf{u})_{K}(\mathbf{W}, \mathbf{v})_{K}=(t \wedge s)(Q \mathbf{u}, \mathbf{v})_{X}$, $\forall s, t \geq 0, \mathbf{u}, \mathbf{v} \in X,(\cdot, \cdot)_{X}$ denotes the inner product of $X, \mathbf{W}(t, \mathbf{x})=\sum_{n=1}^{\infty} \sqrt{\alpha_{n}} \beta_{n}(t) e_{n}(\mathbf{x})$, where $t \wedge$ $s=\min \{t, s\}$ and the Hilbert-Schmidt operator $Q$ is a positive definite, nuclear, symmetric, self-adjoint operator having a finite trace $\operatorname{tr} Q \triangleq \sum_{n=1}^{\infty} \alpha_{n}<+\infty$ with eigenvalues $\alpha_{n},\left\{e_{n}\right\}_{n=1}^{\infty}$ is an orthogonal basis of $X$, and $\left\{\beta_{n}\right\}_{n=1}^{\infty}$ is a sequence of mutually independent standard Brownian motions in $(\Omega, \mathscr{F}, \mathbb{P})($ see $[2,14])$.
(iii) $\mathscr{L}_{2}\left(X_{0}, U\right)$ is the space of all Hilbert-Schmidt operators from $X_{0} \triangleq Q^{1 / 2}(X)$ into $U$; when equipped with the norm $\|\Phi\|_{2} \triangleq \sqrt{\operatorname{tr}\left(\Phi Q \Phi^{*}\right)}$ it becomes a Hilbert space, where $\Phi \in \mathscr{L}_{2}\left(X_{0}, U\right)$; $\Phi^{*}$ denotes the adjoint of $\Phi$.

## 2. Preliminaries and Notations

In this paper, we introduce the following Hilbert spaces: $U=$ $\left\{L^{2}(\mathcal{O})\right\}^{n}$ and $V=\left\{H^{1}(\mathcal{O})\right\}^{n}$; according to [14, 19-21], $V \subset$ $U=U^{\prime} \subset V^{\prime} ; U^{\prime}, V^{\prime}$ denote the dual of the spaces $U$, $V$ respectively, the injection continues, and the embedding is compact; $\|\cdot\|,\| \| \|$ represent the usual norm in $U, V$, respectively. Let us define the operator as follows:

$$
\begin{align*}
A: \Pi(A) & \in U \longrightarrow U \\
A \mathbf{u} & =\nabla \cdot(\mathbf{G} \circ \nabla \mathbf{u})-\mathbf{B u} \tag{3}
\end{align*}
$$

and $\Pi(A)$ is the domain of $A$ defined as $\Pi(A)=\left\{H^{2}(\mathcal{O})\right\}^{n} \cap$ $\left\{H_{0}^{1}(\mathcal{O})\right\}^{n} . A$ is the infinitesimal generator of an analytic semigroup $T(t)$.

Defining the Nemytskii operator as follows:

$$
\begin{equation*}
\mathbf{f}(\mathbf{u})(\mathbf{x})=\tilde{f}(\mathbf{u}(\mathbf{x})), \quad \forall \mathbf{x} \in \mathcal{O} \tag{4}
\end{equation*}
$$

With these notations, we rewrite system (2) in the more abstract form

$$
\begin{align*}
\mathrm{d} \mathbf{u} & =(A \mathbf{u}+\mathbf{C} \mathbf{f}(\mathbf{u})+\mathbf{I}) \mathrm{d} t+\sigma \mathrm{d} \mathbf{W}  \tag{5}\\
\mathbf{u}\left(t_{0}\right) & =\phi
\end{align*}
$$

We recall that

$$
\begin{equation*}
\mathbf{W}_{\mathbf{A}}(t)=\sigma \int_{-\infty}^{t} e^{A(t-s)} \mathrm{d} \mathbf{W}(s) \tag{6}
\end{equation*}
$$

is the solution of the Ornstein-Uhlenbeck process

$$
\begin{equation*}
\mathrm{d} \mathbf{u}=A \mathbf{u} t+\sigma \mathrm{d} \mathbf{W} \tag{7}
\end{equation*}
$$

The regularity of (6) has been proved in $[14,22]$ and $\mathbf{W}_{\mathbf{A}}(t)$ has a $\alpha$-Hölder continuous version with respect to $t, \alpha<1 / 4$. Furthermore, by the law of large number

$$
\begin{align*}
\lim _{t \rightarrow \pm \infty} \frac{\left\|\mathbf{W}_{\mathbf{A}}(t)\right\|}{|t|} & =0 \\
\lim _{t \rightarrow \pm \infty} \frac{\left\|\ln \mathbf{W}_{\mathbf{A}}(t)\right\|}{|t|} & =0 \tag{8}
\end{align*}
$$

(H1) We assume that $\|\mathbf{f}(\mathbf{u})\| \leq k_{1}+k_{2}\|\mathbf{u}\|,\|\mathbf{f}(\mathbf{u})-\mathbf{f}(\mathbf{v})\| \leq$ $k_{3}\|\mathbf{u}-\mathbf{v}\|, \forall \mathbf{u}, \mathbf{v} \in U$.
(H2) We assume that there exists a positive number $\beta$ (O) such that the following Poincaré inequality is valid:

$$
\begin{equation*}
\|\mathbf{u}\| \leq \beta^{-1}\| \| \mathbf{u} \|, \quad \forall \mathbf{u} \in V \tag{9}
\end{equation*}
$$

(H3) We assume that there exists $\alpha>0$ such that $G_{i j}(\mathbf{x}) \geq$ $\alpha / n l, i=1,2, \ldots, n$.
(H4) Let $k_{4}>\sqrt{2} k_{2} k_{5}$, with $k_{4}=\min \left\{b_{1}, b_{2}, \ldots, b_{n}\right\}$ and $k_{5}=\max \left\{\left|c_{i j}\right|\right\}$.

We also need the following propositions in the following sections.

Proposition 1. Consider the following equation:

$$
\begin{align*}
\frac{\mathrm{d}}{\mathrm{~d} t} \mathbf{u} & =A \mathbf{u}, \quad t \geq 0  \tag{10}\\
\mathbf{u}(0) & =\phi
\end{align*}
$$

Suppose (H2), (H3) hold; let $\mathbf{u}(t)=T(t) \boldsymbol{\phi}$ denote the mild solution of (10); then $T(t)$ is a contraction map in $U$.

Proof. We recall that the solution of this linear equation is $\mathbf{u}(t)=e^{A t} \boldsymbol{\phi}$, so $T(t)=e^{A t}$. Now we take the inner product of (10) with $\mathbf{u}(t)$ in $U$, by employing the Gaussian theorem and condition (H3), we get

$$
\begin{equation*}
(A \mathbf{u}, \mathbf{u}) \leq-\alpha\|\mathbf{u}\|^{2}-k_{4}\|\mathbf{u}\|^{2}, \quad \forall \mathbf{u} \in V \tag{11}
\end{equation*}
$$

where $(\cdot, \cdot)$ is the inner product in $U$ (see [1-3]), and we also have

$$
\begin{equation*}
\frac{1}{2} \frac{\mathrm{~d}}{\mathrm{~d} t}\|\mathbf{u}(t)\|^{2}+\alpha\|\mathbf{u}(t)\|^{2}+k_{4}\|\mathbf{u}\|^{2} \leq 0 \tag{12}
\end{equation*}
$$

By (H2), one obtains

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{~d} t}\|\mathbf{u}(t)\|^{2}+2\left(\alpha \beta^{2}+k_{4}\right)\|\mathbf{u}(t)\|^{2} \leq 0 \tag{13}
\end{equation*}
$$

By Gronwall-Bellman inequality, we have

$$
\begin{equation*}
\|\mathbf{u}(t)\|^{2} \leq e^{-2\left(\alpha \beta^{2}+k_{4}\right) t}\|\boldsymbol{\phi}\|^{2} \tag{14}
\end{equation*}
$$

by the definition of $\|T(t)\|$ and uniform boundedness principle, we have

$$
\begin{equation*}
\|T(t)\| \leq e^{-\left(\alpha \beta^{2}+k_{4}\right) t} \leq 1 \tag{15}
\end{equation*}
$$

So $T(t)$ is a contraction map.
Remark 2 (see [14]). Let $T(t)$ be a strongly continuous semigroup and $A$ be the generator of $T(t)$; then we have

$$
\begin{align*}
T\left(t_{1}\right) \phi-T\left(t_{2}\right) \phi & =\int_{t_{2}}^{t_{1}} T(s) A \phi \mathrm{~d} s \\
& =\int_{t_{2}}^{t_{1}} A T(s) \phi \mathrm{d} s, \quad \forall \phi \in \Pi(A) \tag{16}
\end{align*}
$$

Proposition 3 (see [2]). Let $\Psi=\left(\psi_{i j}\right)_{n \times n}$ and $\mathbf{u}=$ $\left(u_{1}, u_{2}, \ldots, u_{n}\right)^{T}$; then $|\Psi \mathbf{u}|_{\mathbb{R}^{n}} \leq \sqrt{n}\|\Psi\|_{F}\|\mathbf{u}\|_{\mathbb{R}^{n}}$, where $\|\mathbf{u}\|_{\mathbb{R}^{n}}$ represents the usual norm of $\mathbb{R}^{n}$ and $\|\Psi\|_{F}$ represents the Frobenius norm of a matrix $\Psi \in \mathbb{R}^{n \times n}$; that is, $\|\Psi\|_{F}=$ $\left(\sum_{i=1}^{n} \sum_{j=1}^{n} \psi_{i j}^{2}\right)^{1 / 2}$.

Let $(X ; d)$ be a complete separable metric space. We shall recall the notions of random dynamical system and random attractor.

Definition 4 (see [22]). Let ( $X, d$ ) be a complete separable metric space. A metric dynamical system (MDS) $\theta \triangleq$ $\left(\Omega, \mathscr{F}, \mathbb{P},\left\{\theta_{t}, t \in R\right\}\right)$ is a family of measure-preserving transformations $\theta_{t}: \Omega \rightarrow \Omega, t \in \mathbb{R}$ such that
(i) $\theta_{0}=i d, \theta_{t} \cdot \theta_{s}=\theta_{t+s}$ for all $t, s \in \mathbb{R}$; • is the composition operator;
(ii) the map $(t, \omega) \rightarrow \theta_{t}(\omega)$ is $\mathscr{B}(\mathbb{R}) \otimes \mathscr{F}$ measurable, and $\theta_{t} \mathbb{P}=\mathbb{P}$ for all $t \in \mathbb{R}$, where $\mathscr{B}(\mathbb{R})$ is the Borel $\sigma$ field of $\mathbb{R}$.

Definition 5 (see [22-25]). A random dynamical system (RDS) over $\theta_{t}$ is a measurable map $\boldsymbol{\varphi}$, such that
(1) $\boldsymbol{\varphi}(0, \omega) x=x, x \in X, \omega \in \Omega$;
(2) $\boldsymbol{\varphi}(s+t, \omega) x=\boldsymbol{\varphi}\left(s, \theta_{t} \omega\right) \cdot \boldsymbol{\varphi}(t, \omega) x, s, t \in \mathbb{R}, x \in X$, $\omega \in \Omega$;
(3) $(t, x) \rightarrow \boldsymbol{\varphi}(t, \omega) x$ is continuous a.e. $\omega \in \Omega$;
(4) $\omega \rightarrow \boldsymbol{\varphi}(t, \omega) x$ is $\mathscr{F}$-measurable for all $(t, x) \in \mathbb{R} \times X$.

Definition 6 (see [26, 27]). A stochastic flow is a family of mappings $S(t, s ; \omega): X \rightarrow X,-\infty<s \leq t<\infty$, parameterized by $\omega$, such that

$$
\begin{equation*}
(t, s, x, \omega) \longrightarrow S(t, s ; \omega) x \tag{17}
\end{equation*}
$$

is $\mathscr{B}(\mathbb{R}) \otimes \mathscr{B}(\mathbb{R}) \otimes \mathscr{B}(X) \otimes \mathscr{F}$-measurable and

$$
\begin{align*}
S(t, r ; \omega) \cdot S(r, s ; \omega) x & =S(t, s ; \omega) x \\
s \leq r \leq t, x & \in X  \tag{18}\\
S(t, s ; \omega) x & =S\left(t-s, 0 ; \theta_{s} \omega\right) x, \quad \omega \in \Omega
\end{align*}
$$

$S$ is said to be a continuous stochastic flow, if $x \rightarrow S(t, s ; \omega) x$ is continuous.

Definition 7 (see [28, 29]). A map $K: \Omega \rightarrow 2^{X}$ is said to be a closed random set if $K(\omega)$ is closed for a.e. $\omega \in \Omega$ and the $\operatorname{map} \omega \rightarrow \operatorname{dist}(x, K(\omega))$ is a.e. measurable for all $x \in X$; dist denotes the Hausdorff semidistance defined as $\operatorname{dist}(x, \mathbb{A})=$ $\inf _{y \in \mathbb{A}} d(x, y)$ and $\operatorname{dist}(\mathbb{A}, \mathbb{B})=\sup _{x \in \mathbb{A}} d(x, \mathbb{B}), \mathbb{A}, \mathbb{B} \in 2^{X}$.

Definition 8 (see $[29,30]$ ). A map $K: \Omega \rightarrow 2^{X}$ is said to be a compact random set if $K(\omega)$ is compact for a.e. $\omega \in \Omega$ and the $\operatorname{map} \omega \rightarrow \operatorname{dist}(x, K(\omega))$ is a.s. measurable.

Definition 9 (see [31, 32]). A random set $K(\omega)$ is called an absorbing set in $X$, if for all $\mathbb{B} \in X$ and a.e. $\omega \in \Omega$, there exists $t_{\mathbb{B}}(\omega)>0$ such that $t \geq t_{\mathbb{B}}(\omega)>0$

$$
\begin{equation*}
\boldsymbol{\varphi}\left(t, \theta_{-t} \omega\right) \mathbb{B} \subset K(\omega) \tag{19}
\end{equation*}
$$

Lemma 10 (Schauder fixed point theorem [33-38]). If $\mathbb{B}$ is a closed bounded convex subset of a space $X$ and $\mathfrak{L}: \mathbb{B} \rightarrow \mathbb{B}$ is completely continuous, then $\mathfrak{L}$ has a fixed point in $\mathbb{B}$.

Lemma 11 (see $[39,40]$ ). Suppose $\varphi$ is a RDS on a Polish space $X$, and there exists a compact set $K(\omega)$ absorbing every bounded deterministic set $\mathbb{B} \in X$. Then the set

$$
\begin{equation*}
\mathfrak{A}(\omega)=\overline{\bigcup_{\mathbb{B} \in X} \bigcap_{T \geq 0} \overline{\bigcup_{t \geq T} \phi\left(t, \theta_{-t} \omega\right) \mathbb{B}}} \tag{20}
\end{equation*}
$$

is a random attractor for $\boldsymbol{\varphi}$.

## 3. Existence and Uniqueness of the Solutions

Let

$$
\begin{equation*}
\mathbf{v}(t)=\mathbf{u}(t)-\mathbf{W}_{\mathbf{A}}(t) \tag{21}
\end{equation*}
$$

where $\mathbf{W}_{\mathbf{A}}(t)$ has been defined in the previous section. Then, from (5) and (6), $\mathbf{v}(t)$ satisfies the equation

$$
\begin{align*}
\mathrm{d} \mathbf{v} & =\left(\mathbf{A} \mathbf{v}+\mathbf{C} \mathbf{f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)+\mathbf{I}\right) \mathrm{d} t  \tag{22}\\
\mathbf{v}\left(t_{0}\right) & =\psi
\end{align*}
$$

where $\psi=\boldsymbol{\phi}-\mathbf{W}_{\mathbf{A}}\left(t_{0}\right)$. Let us rewrite (23) in the integral form

$$
\begin{equation*}
\mathbf{v}(t, \omega)=T(t) \psi+\int_{t_{0}}^{t} T(t-s)\left(\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)+\mathbf{I}\right) \mathrm{d} s \tag{23}
\end{equation*}
$$

Definition 12. If $\mathbf{v}$ satisfies (23), we say that the $\mathbf{u}(t)$ is a mild solution of (1).

Let $X_{T^{*}} \triangleq C\left(\left[t_{0}, t_{0}+T^{*}\right] ; U\right)$, when equipped with the norm $\|\mathbf{u}\|_{X_{T^{*}}}=\left(\sup _{t \in\left[t_{0}, t_{0}+T^{*}\right]}\|\mathbf{u}(t)\|^{2}\right)^{1 / 2}$, it becomes a Banach space. Let

$$
\begin{align*}
& \Sigma\left(m, T^{*}\right)=\left\{\mathbf{v} \in C\left(\left[t_{0}, t_{0}+T^{*}\right] ; U\right):\|\mathbf{v}(t)\| \leq m \forall t\right. \\
& \left.\quad \in\left[t_{0}, t_{0}+T^{*}\right]\right\} \tag{24}
\end{align*}
$$

and consider an initial data $\psi$ which is $\mathscr{F}_{0}$-measurable and belong to $U$, a.e. $\omega \in \Omega$. From now on, we are going to discuss mild solution of equation (22) by Schauder fixed point method in the space $X_{T^{*}}$ for some $T^{*}>0$.

Theorem 13. Assume that $\psi$ is $\mathscr{F}_{0}$-measurable and belongs to $U$, a.e. $\omega \in \Omega$. If conditions (H1)-(H3) hold, then for any $\|\psi\|<$ $m$, there exists a stopping time $T^{*}>0$, such that (1) has a mild solution in $\Sigma\left(m, T^{*}\right)$.

Proof. We split our proof into the following steps.
Step $1 . \Sigma\left(m, T^{*}\right)$ is a nonempty closed bounded convex set.
Let $\mathbf{v}_{1}, \mathbf{v}_{2}$ belong to $\Sigma\left(m, T^{*}\right)$; then for $0 \leq \lambda \leq 1$

$$
\begin{align*}
\left\|\lambda \mathbf{v}_{1}+(1-\lambda) \mathbf{v}_{2}\right\|_{X_{T^{*}}} & \leq \lambda\left\|\mathbf{v}_{1}\right\|_{X_{T^{*}}}+(1-\lambda)\left\|\mathbf{v}_{2}\right\|_{X_{T^{*}}}  \tag{25}\\
& \leq m
\end{align*}
$$

so $\lambda \mathbf{v}_{1}+(1-\lambda) \mathbf{v}_{2} \in \Sigma\left(m, T^{*}\right)$; hence $\Sigma\left(m, T^{*}\right)$ is a convex set.
Step 2. $\mathscr{T}$ maps $\Sigma\left(m, T^{*}\right)$ into $\Sigma\left(m, T^{*}\right)$.
Take any $\mathbf{v}$ in $\Sigma\left(m, T^{*}\right)$ and define $\mathbf{z}=\mathscr{T} \mathbf{v}$ by

$$
\begin{equation*}
\mathbf{z}(t)=T(t) \psi+\int_{t_{0}}^{t} T(t-s)\left(\mathbf{C}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)+\mathbf{I}\right) \mathrm{d} s \tag{26}
\end{equation*}
$$

By the triangle inequality

$$
\begin{align*}
\|\mathbf{z}\| \leq & \|T(t) \psi\| \\
& +\left\|\int_{t_{0}}^{t} T(t-s)\left(\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)+\mathbf{I}\right) \mathrm{d} s\right\| . \tag{27}
\end{align*}
$$

By Proposition 1, it follows that

$$
\begin{equation*}
\|\mathbf{z}(t)\| \leq\|\psi\|+\int_{t_{0}}^{t}\left\|\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)+\mathbf{I}\right\| \mathrm{d} s \tag{28}
\end{equation*}
$$

By using Proposition 3 and norm inequality, as well as condition (H1), we have

$$
\begin{align*}
& \left\|\mathbf{C}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right\| \leq \sqrt{n}\|\mathbf{C}\|_{F}\left\|\mathbf{f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right\| \\
& \quad \leq \sqrt{n}\|\mathbf{C}\|_{F}\left(k_{1}+k_{2}\left\|\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right\|\right) \\
& \quad \leq \sqrt{n}\|\mathbf{C}\|_{F}\left(k_{1}+k_{2}\left(\|\mathbf{v}\|+\left\|\mathbf{W}_{\mathbf{A}}\right\|\right)\right)  \tag{29}\\
& \quad \leq \sqrt{n}\|\mathbf{C}\|_{F}\left(k_{1}+k_{2}\|\mathbf{v}\|_{X_{T^{*}}}+k_{2}\left\|\mathbf{W}_{\mathbf{A}}\right\|_{X_{T^{*}}}\right) ;
\end{align*}
$$

therefore

$$
\begin{align*}
\|\mathbf{z}(t)\|_{X_{T^{*}}} \leq & \|\psi\| \\
& +\int_{t_{0}}^{t}\left(k_{5}\left\|\mathbf{W}_{\mathbf{A}}\right\|_{X_{T^{*}}}+k_{6}+k_{7}\|\mathbf{v}\|_{X_{T^{*}}}\right) \mathrm{d} s  \tag{30}\\
\leq & \|\psi\|+\left(k_{6} \mu+k_{7}+k_{8} m\right) t
\end{align*}
$$

where $\mu=\left\|\mathbf{W}_{\mathbf{A}}\right\|_{X_{T^{*}}}, k_{6}=\sqrt{n} k_{2}\|\mathbf{C}\|_{F}, k_{7}=\|\mathbf{I}\|+\sqrt{n}\|\mathbf{C}\|_{F} k_{1}$, and $k_{8}=\sqrt{n} k_{2}\|\mathbf{C}\|_{F}$.

Hence $\|\mathbf{z}(t)\| \leq m$ for all $t \in\left[t_{0}, t_{0}+T^{*}\right]$, provided that

$$
\begin{equation*}
\|\psi\|+\left(k_{6} \mu+k_{7}+k_{8} m\right)\left(t-t_{0}\right)<m . \tag{31}
\end{equation*}
$$

It is clear that, for any $\|\psi\|<m$, there exists a $T^{*}$ satisfying (31).

Step 3. $\mathscr{T}$ is a equicontinuous map of $\Sigma\left(m, T^{*}\right)$.
Let $t_{1}, t_{2}>0$, then by Remark 2 and Propositions 1 and 3, we have

$$
\begin{align*}
& \| \mathscr{T} \mathbf{v}\left(t_{1}\right)-\mathscr{T} \mathbf{v}\left(t_{2}\right) \| \\
&=\left\|\left(T\left(t_{1}\right)-T\left(t_{2}\right)\right) \psi\right\| \\
&+\left\|\int_{t_{2}}^{t_{1}} T(t-s)\left(\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)+\mathbf{I}\right) \mathrm{d} s\right\| \\
& \leq \int_{t_{2}}^{t_{1}}\|T(s) A \psi\| \mathrm{d} s \\
&+\int_{t_{2}}^{t_{1}}\left\|T(t-s)\left(\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)+\mathbf{I}\right)\right\| \mathrm{d} s  \tag{32}\\
& \leq \int_{t_{2}}^{t_{1}}\|A \psi\| \mathrm{d} s+\int_{t_{2}}^{t_{1}}\left(\left\|\mathbf{C} \mathbf{f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right\|+\|\mathbf{I}\|\right) \mathrm{d} s \\
& \leq \int_{t_{2}}^{t_{1}}\|A \psi\| \mathrm{d} s+\left(k_{6} \mu+k_{7}+k_{8} m\right)\left|t_{1}-t_{2}\right| \\
& \leq\left(\|A \psi\|+k_{6} \mu+k_{7}+k_{8} m\right)\left|t_{1}-t_{2}\right|, \\
& \quad \forall v \in \Sigma\left(m, T^{*}\right)
\end{align*}
$$

so $\mathscr{T} \Sigma\left(m, T^{*}\right)$ is an equicontinuous set.

By the Arzelá-Ascoli theorem, for any bounded set $\mathscr{B} \subset$ $\Sigma\left(m, T^{*}\right)$, the closure of $\mathscr{T} \mathscr{B}$ is compact, so $\mathscr{T}$ is a completely continuous map, then by Steps 1-3 and Lemma 10, $\mathscr{T}$ has a fixed point in $\Sigma\left(m, T^{*}\right)$, which is a mild solution of (1).

Theorem 14. Suppose (H1)-(H3) hold, then for any $\|\psi\|<m$, (1) has at most one mild solution in $\left[t_{0}, t_{0}+T^{*}\right]$.

Proof. Let $\mathbf{v}_{1}, \mathbf{v}_{2} \in \Sigma\left(m, T^{*}\right)$ be two solutions of system (23) on $[0 ; T]$ with $\mathbf{v}_{1}(0)=\mathbf{v}_{2}(0)=\psi$, and set $\mathbf{z}_{i}=\mathscr{T} v_{i}, i=1,2$, and $\mathbf{z}=\mathbf{z}_{1}-\mathbf{z}_{2}$. Then

$$
\begin{align*}
& \mathbf{z}(t) \\
& \quad= T(t)\left(\mathbf{v}_{1}(0)-\mathbf{v}_{2}(0)\right)  \tag{33}\\
&+\int_{t_{0}}^{t} T(t-s) \mathbf{C}\left(\mathbf{f}\left(\mathbf{v}_{1}+\mathbf{W}_{\mathbf{A}}\right)-\mathbf{f}\left(\mathbf{v}_{2}+\mathbf{W}_{\mathbf{A}}\right)\right) \mathrm{d} s .
\end{align*}
$$

Following the method in Step 3 of Theorem 13, we have

$$
\begin{equation*}
\|\mathbf{Z}(t)\|_{X_{T^{*}}} \leq \sqrt{n}\|\mathbf{C}\|_{F} k_{3} T^{*}\left\|\mathbf{v}_{1}-\mathbf{v}_{2}\right\|_{X_{T^{*}}} . \tag{34}
\end{equation*}
$$

We take a stopping time $T^{*}$ such that

$$
\begin{equation*}
\sqrt{n}\|\mathbf{C}\|_{F} k_{3} T^{*}<1 \tag{35}
\end{equation*}
$$

which implies that $\mathbf{v}_{1}(t)=\mathbf{v}_{2}(t)$ and $t \in\left[t_{0}, t_{0}+T^{*}\right]$. By the combination of Theorem 13, there is one unique mild solution $\mathbf{u}(t)$ for (1).

Theorem 15. Suppose (H1)-(H3) hold. If $\mathbf{v} \in C\left(\left[t_{0}, t_{0}+T\right] ; U\right)$ is a solution of system (23), then

$$
\begin{equation*}
\|\mathbf{v}(t)\| \leq\|\psi\|+\sqrt{\delta} \tag{36}
\end{equation*}
$$

for some constant $\mu_{\infty}>0$, where $\delta=e^{|\eta| T}\left(\|\mathbf{I}\|^{2}+\sqrt{n}\|\mathbf{C}\|_{F} k_{1}^{2}+\right.$ $\left.2 k_{2}^{2} \sqrt{n}\|\mathbf{C}\|_{F} \mu_{\infty}^{2}\right)$ and $\gamma=2 \alpha \beta^{2}+2 k_{4}-\sqrt{n}\|\mathbf{C}\|_{F}\left(2 k_{2}^{2}+1\right)-1$.

Proof. By employing the method in [41, 42], let $\left\{\boldsymbol{\psi}^{m}\right\}_{m=1}^{\infty}$ be a sequence in $\left\{C^{\infty}(\mathcal{O})\right\}^{n}$ such that

$$
\begin{equation*}
\psi^{m} \longrightarrow \psi \in U \tag{37}
\end{equation*}
$$

Let $\left\{\mathbf{W}_{\mathbf{A}}^{m}\right\}$ be a sequence of regular process such that

$$
\begin{equation*}
\mathbf{W}_{\mathbf{A}}^{m}(t)=\int_{t_{0}}^{t} T(t-s) \mathrm{d} \mathbf{W}^{m}(s) \longrightarrow \mathbf{W}_{\mathbf{A}}(t) \tag{38}
\end{equation*}
$$

in $\left\{C\left(\left[t_{0}, t_{0}+T\right] \times \mathcal{O}\right)\right\}^{n}$ a.e. $\omega \in \Omega$.
Let $\mathbf{v}^{m}$ be the solution of

$$
\begin{align*}
\mathbf{v}^{m}(t)= & T(t) \psi^{m} \\
& +\int_{t_{0}}^{t} T(t-s)\left(\mathbf{C f}\left(\mathbf{v}^{m}+\mathbf{W}_{A}^{m}\right)+\mathbf{I}\right) \mathrm{d} s \tag{39}
\end{align*}
$$

By using the method in the Theorem 13, it is easy to see that $\mathbf{v}^{m}$ does exist on an interval of time $\left[t_{0}, t_{0}+T^{m}\right]$ such that $T^{m} \rightarrow T^{*}$ a.s. and that $\mathbf{v}^{m}$ converges to $\mathbf{v}$ in $C\left(\left[t_{0}, t_{0}+T\right] ; U\right)$ [39]. Moreover $\mathbf{v}^{m}$ is regular and satisfies

$$
\begin{equation*}
\mathrm{d} \mathbf{v}^{m}=\left(A \mathbf{v}^{m}+\mathbf{C} \mathbf{f}\left(\mathbf{v}^{m}+\mathbf{W}_{A}^{m}\right)+\mathbf{I}\right) \mathrm{d} t \tag{40}
\end{equation*}
$$

Taking the inner product of (40) with $\mathbf{v}^{m}$ in $U$ and employing the result of (11), we find that

$$
\begin{align*}
\frac{1}{2} \frac{\mathrm{~d}}{\mathrm{~d} t}\left\|\mathbf{v}^{m}\right\|^{2}= & \left(\mathbf{v}^{m}, \frac{\mathrm{~d} \mathbf{v}^{m}}{\mathrm{~d} t}\right) \\
= & \left(\mathbf{v}^{m}, A \mathbf{v}^{m}+\mathbf{C} \mathbf{f}\left(\mathbf{v}^{m}+\mathbf{W}_{A}^{m}\right)+\mathbf{I}\right)  \tag{41}\\
\leq & -\left(\alpha \beta^{2}+k_{4}\right)\left\|\mathbf{v}^{m}\right\|^{2} \\
& +\left(\mathbf{v}^{m}, \mathbf{C f}\left(\mathbf{v}^{m}+\mathbf{W}_{A}^{m}\right)\right)+\left(\mathbf{v}^{m}, \mathbf{I}\right)
\end{align*}
$$

By the Cauchy-Schwartz inequality and Young inequality, we get

$$
\begin{equation*}
\left(\mathbf{v}^{m}, \mathbf{I}\right) \leq\left\|\mathbf{v}^{m}\right\|\|\mathbf{I}\| \leq \frac{1}{2}\left\|\mathbf{v}^{m}\right\|^{2}+\frac{1}{2}\|\mathbf{I}\|^{2} \tag{42}
\end{equation*}
$$

By using Young inequality and Proposition 3, as well as condition (H1), one obtains

$$
\begin{align*}
& \left(\mathbf{v}^{m}, \mathbf{C} \mathbf{f}\left(\mathbf{v}^{m}+\mathbf{W}_{A}^{m}\right)\right) \leq\left\|\mathbf{v}^{m}\right\|\left\|\mathbf{C} \mathbf{f}\left(\mathbf{v}^{m}+\mathbf{W}_{A}^{m}\right)\right\| \\
& \leq \sqrt{n}\|\mathbf{C}\|_{F}\left\|\mathbf{v}^{m}\right\|\left\|\mathbf{f}\left(\mathbf{v}^{m}+\mathbf{W}_{A}^{m}\right)\right\| \\
& \leq \frac{1}{2} \sqrt{n}\|\mathbf{C}\|_{F}\left(\left\|\mathbf{v}^{m}\right\|^{2}+\left\|\mathbf{f}\left(\mathbf{v}^{m}+W_{A}^{m}\right)\right\|^{2}\right)  \tag{43}\\
& \leq \frac{1}{2} \sqrt{n}\|\mathbf{C}\|_{F}\left(\left\|\mathbf{v}^{m}\right\|^{2}+2 k_{1}^{2}+2 k_{2}^{2}\left\|\mathbf{v}^{m}+\mathbf{W}_{A}^{m}\right\|^{2}\right) \\
& \leq \frac{1}{2} \sqrt{n}\|\mathbf{C}\|_{F}\left(\left(4 k_{2}^{2}+1\right)\left\|\mathbf{v}^{m}\right\|^{2}+2 k_{1}^{2}+4 k_{2}^{2}\left\|\mathbf{W}_{A}^{m}\right\|^{2}\right)
\end{align*}
$$

By (H2) and (H3), we deduce that

$$
\begin{align*}
\frac{\mathrm{d}}{\mathrm{~d} t} & \left\|\mathbf{v}^{m}\right\|^{2} \\
\leq & -2\left(\alpha \beta^{2}+k_{4}\right)\left\|\mathbf{v}^{m}\right\|^{2}  \tag{44}\\
& +\sqrt{n}\|\mathbf{C}\|_{F}\left(\left(4 k_{2}+1\right)\left\|\mathbf{v}^{m}\right\|^{2}+2 k_{1}^{2}+4 k_{2}^{2}\left\|\mathbf{W}_{A}^{m}\right\|^{2}\right) \\
& +\left\|\mathbf{v}^{m}\right\|^{2}+\|\mathbf{I}\|^{2} \leq-\eta\left\|\mathbf{v}^{m}\right\|^{2}+\kappa
\end{align*}
$$

where $\gamma=2 \alpha \beta^{2}+2 k_{4}-\sqrt{n}\|\mathbf{C}\|_{F}\left(4 k_{2}^{2}+1\right)-1$ and $\kappa=\|\mathbf{I}\|^{2}+$ $2 \sqrt{n}\|\mathbf{C}\|_{F} k_{1}^{2}+4 k_{2}^{2} \sqrt{n}\|\mathbf{C}\|_{F}\left\|\mathbf{W}_{A}^{m}\right\|^{2}$. By the classical Gronwall inequality, then we have

$$
\begin{aligned}
& \left\|\mathbf{v}^{m}(t)\right\|^{2} \leq e^{-\gamma t}\left\|\psi^{m}\right\|^{2}+\int_{t_{0}}^{t} e^{-\gamma s} \kappa \mathrm{~d} s \leq\left\|\boldsymbol{\psi}^{m}\right\|^{2} \\
& \quad+e^{|\gamma| T}\left(\|\mathbf{I}\|^{2}+2 \sqrt{n}\|\mathbf{C}\|_{F} k_{1}^{2}+4 k_{2}^{2} \sqrt{n}\|\mathbf{C}\|_{F} \mu_{m, \infty}^{2}\right) \\
& \quad \cdot T
\end{aligned}
$$

with $\mu_{m, \infty}=\sup _{t \in\left[t_{0}, t_{0}+T\right]}\left\|\mathbf{W}_{A}^{m}(t)\right\|$, for a.e. $\omega \in \Omega$.
Taking the limit as $m \rightarrow \infty$, we see that a.s.

$$
\begin{equation*}
\|\mathbf{v}(t)\|^{2} \leq\|\psi\|^{2}+\delta . \tag{46}
\end{equation*}
$$

It follows that

$$
\begin{equation*}
\|\mathbf{v}(t)\| \leq\|\psi\|+\sqrt{\delta} \tag{47}
\end{equation*}
$$

thus we complete the proof.

It is easy to derive the following from Theorems 13, 14, and 15.

Theorem 16. Let $\phi$ be given which is $\mathscr{F}_{0}$-measurable and bounded in $U$, a.s.; then there exists a unique mild solution $\mathbf{u}$ of (2), with $\mathbf{u} \in C\left(\left[t_{0}, t_{0}+T\right] ; U\right), \forall T<\infty$.

## 4. Existence of the Random Attractor

We refer $\omega=\mathbf{W}(t)$ and $\Omega=\left\{\omega \in C\left(\mathbb{R}, \mathbb{R}^{n}\right) \mid \boldsymbol{\omega}(0)=\mathbf{0}\right\}$, with $\mathbb{P}$ being the product measure of two Wiener measures on the negative and positive parts of $\Omega$. In this case, there exists a Wiener shift $\theta_{t}$ defined as $\theta_{t} \omega(s)=\omega(t+s)-\omega(s), s, t \in \mathbb{R}$, which is an ergodic transformation.

Let $\mathbf{v}\left(t, \omega ; t_{0}, \psi\right)$ be the solution of (23) with $\mathbf{v}(0)=\psi$ and $\mathbf{u}\left(t, \omega ; t_{0}, \phi\right)$ is the mild solution of (1) with $\mathbf{u}(0)=\boldsymbol{\phi}$. Then, from Theorem 15, we know that the map $\psi \rightarrow \mathbf{v}\left(t, \omega ; t_{0}, \psi\right)$ is continuous, by $\mathbf{v}(t)=\mathbf{u}(t)-\mathbf{W}_{A}(t)$; then one can define the maps $S(t, s ; \omega)$ and $\boldsymbol{\varphi}(t, \omega)$ by

$$
\begin{align*}
S\left(t, t_{0} ; \omega\right) \boldsymbol{\phi} & =\mathbf{v}\left(t, \omega ; t_{0}, \psi\right)+\mathbf{W}_{\mathbf{A}}(t, \omega)  \tag{48}\\
\boldsymbol{\varphi}(t, \omega) \boldsymbol{\phi} & =S(t, 0 ; \omega) \boldsymbol{\phi}=\mathbf{v}(t, 0 ; \omega) \psi+\mathbf{W}_{\mathbf{A}}(t, \omega) .
\end{align*}
$$

By the result of Theorem 15, $S$ is a continuous stochastic flow and $\varphi$ is a continuous RDS associated with (1).
4.1. Absorbing Sets in $U$. Let us define the operator as follows:

$$
\begin{align*}
\mathscr{A}: \Pi(\mathscr{A}) & \in U \longrightarrow U \\
\mathscr{A} \mathbf{u} & =-\nabla \cdot(G(\mathbf{x}) \circ \nabla \mathbf{u}) \tag{49}
\end{align*}
$$

and $\Pi(\mathscr{A})=\Pi(A)=\left\{H^{2}(\mathcal{O})\right\}^{n} \cap\left\{H_{0}^{1}(\mathcal{O})\right\}^{n}$. We can infer from (10) and (49) that

$$
\begin{equation*}
A \mathbf{u}=-\mathscr{A} \mathbf{u}-\mathbf{B} \mathbf{u} . \tag{50}
\end{equation*}
$$

The associated bilinear operator with $\mathscr{A}$ is defined as $a(\mathbf{u}, \mathbf{v})=$ $(\mathscr{A} \mathbf{u}, \mathbf{v})$. By utilizing the result of [2], we have

$$
\begin{equation*}
a(\mathbf{u}, \mathbf{u})=\int_{\mathscr{O}} G:(\nabla \mathbf{u} \circ \nabla \mathbf{u}) \mathrm{d} x \geq \alpha\|\mathbf{u}\|^{2} \tag{51}
\end{equation*}
$$

where $\mathbf{A}: \mathbf{B}$ is the Frobenius inner product of two $n \times m$ matrixes, defined as $\mathbf{A}: \mathbf{B} \triangleq \sum_{i}^{n} \sum_{j}^{m} a_{i j} b_{i j}$.

Let $t_{0}<-1$ and $\psi \in U$ be given, and let $\mathbf{v}$ be the solution of equation (23). Taking the inner product of (22) with $\mathbf{v}$ in $U$ we find that

$$
\begin{align*}
\frac{1}{2} \frac{\mathrm{~d}}{\mathrm{~d} t}\|\mathbf{v}\|^{2}= & \left(\mathbf{v}, \frac{\mathrm{d} \mathbf{v}}{\mathrm{~d} t}\right)=(\mathbf{v}, A \mathbf{v})+\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)+\mathbf{I} \\
\leq & -a(\mathbf{v}, \mathbf{v})-k_{4}\|\mathbf{v}\|^{2}+\left(\mathbf{v}, \mathbf{C}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right)  \tag{52}\\
& +(\mathbf{v}, \mathbf{I}) .
\end{align*}
$$

By the Cauchy-Schwartz inequality and Young inequality

$$
\begin{equation*}
(\mathbf{v}, \mathbf{I}) \leq\|\mathbf{v}\|\|\mathbf{I}\| \leq \frac{k_{4}}{4}\|\mathbf{v}\|^{2}+\frac{1}{k_{4}}\|\mathbf{I}\|^{2} . \tag{53}
\end{equation*}
$$

By using Young inequality and (H1)

$$
\begin{align*}
\left(\mathbf{v}, \mathbf{C} \mathbf{f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right) \leq & \leq \mathbf{v}\| \| \mathbf{C}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right) \| \\
& \leq k_{5}\|\mathbf{v}\|\left\|\mathbf{f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right\| \\
\leq & \frac{k_{4}}{4}\|\mathbf{v}\|^{2}+\frac{k_{5}^{2}}{k_{4}}\left\|\mathbf{f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right\|^{2}  \tag{54}\\
\leq & \left(\frac{k_{4}}{4}+\frac{4 k_{2}^{2} k_{5}^{2}}{k_{4}}\right)\|\mathbf{v}\|^{2} \\
& +\frac{2 k_{5}^{2}}{k_{4}}\left(k_{1}^{2}+2 k_{2}^{2}\left\|\mathbf{W}_{\mathbf{A}}\right\|^{2}\right)
\end{align*}
$$

We deduce from (51)-(54) that

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{~d} t}\|\mathbf{v}\|^{2}+2 a(\mathbf{v}, \mathbf{v}) \leq-\eta\|\mathbf{v}\|^{2}+P_{1}(t, \omega) \tag{55}
\end{equation*}
$$

where $\eta=k_{4}-2 k_{2}^{2} k_{5}^{2} / k_{4}>0$ and $P_{1}(t, \omega)=\left(2 k_{5}^{2} / k_{4}\right)\left(k_{1}+\right.$ $\left.2 k_{2}\left\|\mathbf{W}_{\mathbf{A}}\right\|^{2}\right)+\left(2 / k_{4}\right)\|\mathbf{I}\|^{2}$. By (H4), we know that $\eta>0$; using the classical Gronwall inequality, for $t_{0} \leq-1$, we have

$$
\begin{align*}
\|\mathbf{v}(-1)\|^{2} \leq & e^{-\eta\left(-1-t_{0}\right)}\left\|\mathbf{v}\left(t_{0}\right)\right\|^{2} \\
& +\int_{t_{0}}^{-1} e^{\eta(-1-s)} P_{1}(s, \omega) \mathrm{d} s \\
= & e^{-\eta\left(-1-t_{0}\right)}\left\|\mathbf{u}\left(t_{0}\right)\right\|^{2}  \tag{56}\\
& +e^{-\eta\left(-1-t_{0}\right)}\left\|\mathbf{W}_{\mathbf{A}}\left(t_{0}\right)\right\|^{2} \\
& +\int_{t_{0}}^{-1} e^{-\eta(-1-s)} P_{1}(s, \omega) \mathrm{d} s
\end{align*}
$$

Theorem 17. Under conditions (H1)-(H4), the RDS $\varphi$ defined by (48) admits a random absorbing set in $U$.

Proof. Let the random variable

$$
\begin{equation*}
r_{1}^{2}(\omega)=2+\int_{-\infty}^{-1} e^{-\eta(-1-s)} P_{1}(s, \omega) \mathrm{d} s \tag{57}
\end{equation*}
$$

By (8), we know that $P_{1}(s, \omega)$ has at most polynomial growth for $s \rightarrow-\infty$, so $r(\omega)$ is finite. On the other hand, $\left\|\mathbf{W}_{\mathbf{A}}\left(t_{0}\right)\right\|^{2}$ also has at most polynomial growth for $t_{0}$, as $t_{0}$ tending to $-\infty$, according to (8). So, we can choose $\bar{t}$ such that

$$
\begin{equation*}
e^{-\eta\left(-1-t_{0}\right)}\left(\|\phi\|^{2}+\left\|\mathbf{W}_{\mathrm{A}}\left(t_{0}\right)\right\|^{2}\right) \leq 2, \quad t_{0} \leq \bar{t} \tag{58}
\end{equation*}
$$

By (56), (57), and (58), for all $\rho>0$, there exists $t(\omega) \leq-1$, such that for all $t_{0} \leq t(\omega)$ and $\phi \in U$, with $\|\boldsymbol{\phi}\| \leq \rho$, the following inequality is satisfied:

$$
\begin{equation*}
\left\|\mathbf{v}\left(-1, \omega ; t_{0}, \psi\right)\right\|^{2} \leq r_{1}^{2}(\omega) \tag{59}
\end{equation*}
$$

By (48) and (59), the stochastic flow of system (1) satisfies

$$
\begin{equation*}
\left\|S\left(-1, t_{0} ; \omega\right) \phi\right\| \leq r_{1}(\omega)+\left\|\mathbf{W}_{\mathbf{A}}\left(t_{0}, \omega\right)\right\| \triangleq r_{2}(\omega) \tag{60}
\end{equation*}
$$

By using the relationship between stochastic flow and RDS

$$
\begin{align*}
\boldsymbol{\varphi} & \left(t, \theta_{-t} \omega\right) B_{U}(\mathbf{0}, \rho)=S\left(t, 0 ; \theta_{-t} \omega\right) B_{U}(\mathbf{0}, \rho) \\
& =S(0,-t ; \omega) B_{U}(\mathbf{0}, \rho)  \tag{61}\\
& =S(0,-1 ; \omega) S(-1,-t ; \omega) B_{U}(\mathbf{0}, \rho) \\
& \subset S(0,-1 ; \omega) B_{U}\left(\mathbf{0}, r_{2}(\omega)\right), \quad \forall \rho>0
\end{align*}
$$

where $B_{U}(\mathbf{0}, \rho)$ denotes the ball of $U$ centered at $\mathbf{0}$ with radius $\rho$, let $K_{1}(\omega)=S(0,-1 ; \omega) B_{U}\left(\mathbf{0}, r_{2}(\omega)\right)$; there exists $t_{B}(\omega)>0$ such that $t \geq t_{B}(\omega)>0$

$$
\begin{equation*}
\boldsymbol{\varphi}\left(t, \theta_{-t} \omega\right) B_{U}(\mathbf{0}, \rho) \subset K_{1}(\omega), \quad \text { a.e. } \omega \in \Omega \tag{62}
\end{equation*}
$$

By Definition 9, we get the conclusion.
We need the following auxiliary proposition in the following sections.

Theorem 18. There exists a random variable $r_{3}(\omega), r_{4}(\omega)>0$, such that, for all $\rho>0$, there exist $t(\omega) \leq-1$, such that for all $t_{0} \leq t(\omega)$ and all $\phi \in U$, with $\|\phi\| \leq \rho$, the solution $\mathbf{v}\left(t, \omega ; t_{0}, \psi\right)$ of system (22) satisfies the inequality

$$
\begin{gather*}
\int_{-1}^{0}\left\|\mathbf{v}\left(s, \omega ; t_{0}, \psi\right)\right\|^{2} \mathrm{~d} s \leq r_{3}(\omega),  \tag{63}\\
\int_{-1}^{0}\|a(\mathbf{v}, \mathbf{v})\| \mathrm{d} s \leq r_{4}(\omega) .
\end{gather*}
$$

Proof. From (55), we can get

$$
\begin{equation*}
\int_{-1}^{0} a(\mathbf{v}, \mathbf{v}) \mathrm{d} s \leq \frac{1}{2}\|\mathbf{v}(-1)\|^{2}+\frac{1}{2} \int_{-1}^{0} P_{1}(s, \omega) \mathrm{d} s \tag{64}
\end{equation*}
$$

Integrating (55) over $[-1,0]$, we can also get

$$
\begin{equation*}
\int_{-1}^{0}\|\mathbf{v}\|^{2} \mathrm{~d} s \leq \frac{1}{\eta}\|\mathbf{v}(-1)\|^{2}+\frac{1}{\eta} \int_{-1}^{0} P_{1}(s, \omega) \mathrm{d} s \tag{65}
\end{equation*}
$$

Let $r_{3}(\omega)=(1 / 2)\|\mathbf{v}(-1)\|^{2}+(1 / 2) \int_{-1}^{0} P_{1}(s, \omega) \mathrm{d} s, r_{4}(\omega)=$ $(1 / \eta)\|\mathbf{v}(-1)\|^{2}+(1 / \eta) \int_{-1}^{0} P_{1}(s, \omega) \mathrm{d} s$, by Theorem $17,\|\mathbf{v}(-1)\|$ is finite, and by using (8) again, we know that $r_{3}(\omega), r_{4}(\omega)$ is finite.
4.2. Absorbing Set in $V$. We now prove the existence of an absorbing set in $V$. Multiplying (22) by $\mathscr{A} \mathbf{v}$ and integrating over $\mathcal{O}$, by (50), we have

$$
\begin{align*}
\left(\mathscr{A} \mathbf{v}, \frac{\mathrm{d} \mathbf{v}}{\mathrm{~d} t}\right)= & \left(\mathscr{A} \mathbf{v}, A \mathbf{v}+C \mathbf{f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)+\mathbf{I}\right) \\
\leq & -\|\mathscr{A} \mathbf{v}\|^{2}-(\mathscr{A} \mathbf{v}, \mathbf{B} \mathbf{v})  \tag{66}\\
& +\|\mathscr{A} \mathbf{v}\|\left\|\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right\|+(\mathscr{A} \mathbf{v}, \mathbf{I}) .
\end{align*}
$$

Using the Dirichlet boundary condition and the Green first identity

$$
\begin{equation*}
\left(\mathscr{A} \mathbf{v}, \frac{\mathrm{d} \mathbf{v}}{\mathrm{~d} t}\right)=a\left(\mathbf{v}, \frac{\mathrm{~d}}{\mathrm{~d} t} \mathbf{v}\right)=\frac{1}{2} \frac{\mathrm{~d}}{\mathrm{~d} t} a(\mathbf{v}, \mathbf{v}) . \tag{67}
\end{equation*}
$$

By Gauss formula [3]

$$
\begin{equation*}
(\mathscr{A} \mathbf{v}, \mathbf{B} \mathbf{v})=\int_{O}(B G):(\nabla \mathbf{v} \circ \nabla \mathbf{v}) \mathrm{d} \mathbf{x} \geq k_{4} \alpha\|\mathbf{v}\|^{2} \tag{68}
\end{equation*}
$$

We can also infer from the divergence theorem

$$
\begin{equation*}
(\mathscr{A} \mathbf{v}, \mathbf{I})=-\sum_{i=1}^{n} \int_{\partial \mathscr{O}} I_{i}\left(G_{i} \nabla u_{i}\right) \cdot \mathrm{d} s=0 . \tag{69}
\end{equation*}
$$

Using the Young inequality again, we get

$$
\begin{equation*}
\|\mathscr{A} \mathbf{v}\|\left\|\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathrm{A}}\right)\right\| \leq\|\mathscr{A} \mathbf{v}\|^{2}+\frac{1}{4}\left\|\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right\|^{2} \tag{70}
\end{equation*}
$$

we can deduce from Proposition 3 and condition (H1)

$$
\begin{align*}
& \left\|\mathbf{C f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right\|^{2} \leq n\|\mathbf{C}\|_{F}^{2}\left\|\mathbf{f}\left(\mathbf{v}+\mathbf{W}_{\mathbf{A}}\right)\right\|^{2} \\
& \quad \leq n\|\mathbf{C}\|_{F}^{2}\left(4 k_{2}^{2}\|\mathbf{v}\|^{2}+2 k_{1}^{2}+4 k_{2}^{2}\left\|\mathbf{W}_{\mathbf{A}}\right\|^{2}\right) . \tag{71}
\end{align*}
$$

It follows from (66)-(71) that

$$
\begin{equation*}
\frac{\mathrm{d}}{\mathrm{~d} t} a(\mathbf{v}, \mathbf{v}) \leq-k_{9} a(\mathbf{v}, \mathbf{v})+P_{2}(t, \omega) \tag{72}
\end{equation*}
$$

where $k_{9}=2 k_{4}$

$$
\begin{equation*}
P_{2}(t, \omega)=n\|\mathbf{C}\|_{F}^{2}\left(2 k_{2}^{2}\|\mathbf{v}\|^{2}+k_{1}^{2}+2 k_{2}^{2}\left\|\mathbf{W}_{\mathbf{A}}\right\|^{2}\right) . \tag{73}
\end{equation*}
$$

Integrating (72) over [ $s, 0$ ], we have

$$
\begin{equation*}
a(\mathbf{v}(0), \mathbf{v}(0)) \leq a(\mathbf{v}(s), \mathbf{v}(s))+\int_{s}^{0} P_{2}(s, \omega) \mathrm{d} s \tag{74}
\end{equation*}
$$

integrating (72) again in $s$ over $[-1,0]$ and using (51), we have

$$
\begin{align*}
\|\mathbf{v}(0)\|^{2} \leq & \frac{1}{\alpha} \int_{-1}^{0} a(\mathbf{v}(s), \mathbf{v}(s)) \mathrm{d} s \\
& +\frac{1}{\alpha} \int_{-1}^{0} P_{2}(s, \omega) \mathrm{d} s \tag{75}
\end{align*}
$$

Theorem 19. Under conditions (H1)-(H4), the RDS $\boldsymbol{\varphi}$ defined by (48) admits an absorbing set in $V$.

Proof. Put the random variable

$$
\begin{equation*}
r_{5}^{2}(\omega) \triangleq \frac{1}{\alpha} \int_{-1}^{0} a(\mathbf{v}(s), \mathbf{v}(s)) \mathrm{d} s+\frac{1}{\alpha} \int_{-1}^{0} P_{2}(s, \omega) \mathrm{d} s \tag{76}
\end{equation*}
$$

By Theorem 18, $r_{5}$ is a finite number; then by employing (71), for all $\rho>0$, there exist $t(\omega) \leq-1$, for all $t_{0} \leq t(\omega)$ and all $\phi \in V$, with $\|\boldsymbol{\phi}\| \leq \rho$; the solution $\mathbf{v}\left(0, \omega ; t_{0}, \psi\right)$ of system (23) satisfies the inequality

$$
\begin{equation*}
\left\|\mathbf{v}\left(0, \omega ; t_{0}, \boldsymbol{\phi}-\mathbf{W}_{\mathbf{A}}(0, \omega)\right)\right\|^{2} \leq r_{5}^{2}(\omega) \tag{77}
\end{equation*}
$$

which also means the stochastic flow of (1) satisfies

$$
\begin{equation*}
\left\|S\left(0, t_{0} ; \omega\right) \boldsymbol{\phi}\right\|\left\|\leq r_{5}(\omega)+\right\| \mathbf{W}_{\mathbf{A}}\left(t_{0}, \omega\right)\| \| r_{6}(\omega) \tag{78}
\end{equation*}
$$



Figure 1: Frequency of Example 1.

By using the definition of RDS and the stochastic flow, we have

$$
\begin{equation*}
\boldsymbol{\varphi}\left(t, \theta_{-t} \omega\right)=S\left(t, 0 ; \theta_{-t} \omega\right)=S(0,-t ; \omega) \tag{79}
\end{equation*}
$$

By (78) and (79), $\forall\|\phi\| \leq \rho$, for sufficiently large $t$

$$
\begin{equation*}
\left\|\boldsymbol{\varphi}\left(t, \theta_{-t} \omega\right) \boldsymbol{\phi}\right\| \leq r_{6}(\omega) \tag{80}
\end{equation*}
$$

Let $K_{2}(\omega)=B_{V}\left(\mathbf{0}, r_{6}(\omega)\right)$, where $B_{V}\left(\mathbf{0}, r_{6}\right)$ denotes the ball of $V$ centered at $\mathbf{0}$ with radius $r_{6}$, so for all $B_{V}(\mathbf{0}, \rho) \in U$, there exists $t_{B}(\omega)>0$ such that $t \geq t_{B}(\omega)>0$

$$
\begin{equation*}
\boldsymbol{\varphi}\left(t, \theta_{-t} \omega\right) B_{V}(0, \rho) \subset K_{2}(\omega), \quad \text { a.e. } \omega \in \Omega \tag{81}
\end{equation*}
$$

Theorem 20. Assume that (H1)-(H4) hold; then the RDS $\boldsymbol{\varphi}$ generated by (1) possesses a random attractor in $U$.

Proof. By Theorem 19, $\boldsymbol{\varphi}(t ; \omega) \boldsymbol{\phi}$ is absorbed by the random set $K_{2}(\omega)$. Since the embedding of $V$ to $U$ is compact, so the bounded set $K_{2}(\omega)$ is compact. A combination of Lemma 11 and Theorems 17 and 19, we get the existence of random attractor for the RDS $\boldsymbol{\varphi}$.

If the active function $\mathbf{f}$ is a bounded Lipschitz function, then we can choose a sufficiently large $k_{1}$ such that $k_{2}=0$, so (H4) is satisfied automatically; then we get the following theorem.

Corollary 21. Under conditions (H1)-(H3) and the fact that there exists a constant $k_{1}$ such that $\|\mathbf{f}(\mathbf{u})\| \leq k_{1}$, the RDS $\boldsymbol{\varphi}$ defined by (48) admits a random attractor in $U$.

## 5. Example and Simulation

Example 1.

$$
\begin{aligned}
\dot{u}_{1}(t, \mathbf{x})= & \Delta u_{1}-2 u_{1}-2 \tanh \left(u_{1}(t-1, x)\right) \\
& -\tanh \left(u_{2}(t-1, x)\right)+1.3 \frac{\partial^{2} W}{\partial t \partial x} \\
\dot{u}_{2}(t, \mathbf{x})= & \Delta u_{2}-2 u_{2}+\tanh \left(u_{1}(t-1, x)\right) \\
& -3 \tanh \left(u_{2}(t-1, x)\right)+1.5 \frac{\partial^{2} W}{\partial t \partial x} \\
u_{i}(t, 0)= & u_{i}(t, 20)=0, \quad t \geq 0, \\
u_{1}(0, \mathbf{x})= & \sin (0.2 \pi \mathbf{x}), \\
u_{2}(0, \mathbf{x})= & \sin (0.2 \pi \mathbf{x}),
\end{aligned}
$$

$$
x \in \mathcal{O}=[0,20] .
$$

Proof. In this case $U=\left\{L^{2}(\mathcal{O})\right\}^{2}$ and $\mathbf{f}(\mathbf{u})$ is a bounded global Lipschitz continuous function with $k_{1}=20, k_{2}=0$, and $k_{3}=1,\|S(t)\| \leq 1$, and $\Delta$ is the Laplace operator. $E W=0$ and $W$ is the cylindrical Wiener process with $E[W(s$, $x) \wedge W(t, y)]=(t \wedge s)(x \wedge y), \forall s, t \geq 0, x, y \in[0,20]$, $E$ is the expectation, and $a \wedge b=\min \{a, b\}$. Under these assumptions, by Theorem 15, this system has a global mild solution. Meanwhile $k_{4} \geq 2 \sqrt{2} k_{2} k_{5}$, so according to Theorem 20, this system possesses a random attractor.

We simulate this example by using the Matlab; for detailed information see Figures 1-3. A Crank-Nicolson method in time and second-order center differences in space are used to discrete this model. A Newton iterative method is used to solve the discretized nonlinear equation. For more theories


Figure 2: Simulation of $u_{1}$ in Example 1.


Figure 3: Simulation of $u_{2}$ in Example 1.
about the numerical theory about SPDE, we refer to [42, 43].

The method used in this article can be extended to other systems, such as the biological systems and fluid mechanical systems [44-48]. We can also apply this method to the system driven by other types of noise, such as the G-Brown motion [49]. By the way, some control technique may be used to stabilize these systems with random attractor [50-54]. We will study these problems in the future.

## Conflicts of Interest

The authors declare that they have no conflicts of interest.

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