# WHITE NOISE INDEXED BY LOOPS 


#### Abstract

By Ognian B. Enchev Boston University Given a Riemannian manifold $M$ and loop $\phi: S^{1} \mapsto M$, we construct a Gaussian random process $S^{1} \ni \theta \leadsto X_{\theta} \in \mathrm{T}_{\phi(\theta)} \mathrm{M}$, which is an analog of the Brownian motion process in the sense that the formal covariant derivative $\theta \leadsto \nabla_{\theta} X_{\theta}$ appears as a stationary process whose spectral measure is uniformly distributed over some discrete set. We show that $X$ satisfies the two-point Markov property (reciprocal process) if the holonomy along the loop $\phi$ is nontrivial. The covariance function of X is calculated and the associated abstract Wiener space is described. We also characterize $X$ as a solution of a special (nondiffusion type) stochastic differential equation. Somewhat surprisingly, the nature of $X$ turns out to be very different if the holonomy along $\phi$ is the identity map I: $\mathrm{T}_{\phi(0)} \mathrm{M} \rightarrow$ $\mathrm{T}_{\phi(0)} \mathrm{M}$. In this case, we show that the usual periodic Ornstein-Uhlenbeck process, associated with a harmonic oscillator at nonzero temperature, may be viewed as a standard velocity process in which the driving Brownian motion is replaced by the process $X$.


1. Overview and motivation. The present paper grew out of the desire to construct a stochastic process $\phi_{\mathrm{t}}, \mathrm{t} \in \mathbf{R}_{+}$, with values in

$$
\mathbf{L}(\mathrm{M}):=\left\{\phi \in \mathrm { C } \left(\left[0,2 \pi[\rightarrow \mathrm{M}) \mid \lim _{\theta \not 2 \pi} \phi(\theta)=\phi(0)\right\},\right.\right.
$$

the space of loops over a Riemannian manifold M, which incorporates as much symmetry as possible More specifically, we want the probability distributions in the cross-section of TM above $\phi$ of the formal differential $\theta \leadsto$ $\mathrm{d} \phi_{\mathrm{t}}(\theta) \equiv \mathrm{d}_{\mathrm{t}} \phi_{\mathrm{t}}(\theta)$ and its (formal) covariant derivative $\theta \leadsto \nabla_{\theta} \mathrm{d} \phi_{\mathrm{t}}(\theta)$ both to be Gaussian and invariant under rotation along the loop, with $\theta \leadsto \nabla_{\theta} \mathrm{d} \phi_{\mathrm{t}}(\theta)$ having uniform spectral density (white noise along the loop $\phi$ ). If parallel translation along $\phi$ is possible, by choosing an orthonormal frame in $\mathrm{T}_{\phi(0)} \mathrm{M}$ and moving it parallel along $\phi$ in positive direction, $\theta \leadsto \mathrm{d} \phi_{\mathrm{t}}(\theta)$ could be treated as a Gaussian process in $\mathbf{R}^{d}$ indexed by $\theta \in \mathbf{R}$, which is stationary and pseudoperiodic, in that its value at $\theta+2 \pi$ is the one at $\theta$ twisted by the holonomy along $\phi$. In terms of Dirichlet forms, processes on $\mathbf{L}(\mathrm{M})$ were constructed in [1] by considering probability distributions in the cross-section of TM above $\phi \in \mathbf{L}(\mathrm{M})$ derived from the abstract Wiener space construction associated with the Hilbert space $H$ of absolutely continuous mappings

[^0]$S^{1} \ni \theta \leadsto \mathrm{~h}(\theta) \in \mathrm{T}_{\phi(\theta)} \mathrm{M}$ endowed with the scalar product
$$
(\mathrm{h} \mid \mathrm{k}):=\frac{1}{2 \pi} \int_{0}^{2 \pi}\langle\mathrm{~h}(\theta), \mathrm{k}(\theta)\rangle \mathrm{d} \theta+\frac{1}{2 \pi} \int_{0}^{2 \pi}\left\langle\nabla_{\theta} \mathrm{h}(\theta), \nabla_{\theta} \mathrm{k}(\theta)\right\rangle \mathrm{d} \theta, \mathrm{~h}, \mathrm{k} \in \mathrm{H},
$$
where $\langle\cdot, \cdot\rangle:=$ the Riemannian metric on $M$. It is easy to check that although this construction yields a rotation-invariant process the spectrum of its (formal) covariant derivative is not flat. In fact, this is precisely the periodic Ornstein-Uhlenbeck process described in Section 4 with $\mathrm{m}=1$.

Alternatively, one may incorporate symmetry in the process $\theta \leadsto \mathrm{d} \phi_{\mathrm{t}}(\theta)$ by requiring it to be a reciprocal process, that is, to share the two-point Markov property, in addition to being rotation-along-the-loop invariant (obviously, the standard Markov property does not make sense for processes indexed by the circle). This approach was taken in the recent work [3], which extends some of the results of [2] on periodic Osterwalder-Schrader positive processes. Unfortunately, if the hol onomy along the loop $\phi$ is trivial, the requirement that $S^{1} \ni \theta \leadsto X_{\theta}$ be two-point Markov contradicts the requirement that $\mathrm{S}^{1} \ni \theta \leadsto \nabla_{\theta} \mathrm{X}_{\theta}$ has a flat spectrum. On the other hand, in the case of nontrivial holonomy, the flatness of the spectrum of $S^{1} \ni \theta \rightsquigarrow \nabla_{\theta} X_{\theta}$ entails that $\mathrm{S}^{1} \ni \theta \leadsto \mathrm{X}_{\theta}$ is two-point Markov. These statements, which we establish in Section 3, have two interesting consequences. First, while one would expect "white noise" to be a much more restrictive property than "Markov," this is not always the case; in fact, somewhat surprisingly, it fails to be the case in the simplest possible scenario: loops with trivial hol onomy. The second observation is that replacing the index space $\mathbf{R}_{+}$by some loop $\phi \in \mathbf{L}(\mathrm{M})$ makes the Markovian nature of the white noise process more intriguing. So, we treat the case $\operatorname{Holonomy}(\phi)=$ Id separately from the case $\operatorname{Holonomy}(\phi) \neq \mathrm{Id}$ and in each case give a complete description of the covariance nature and of the Markovian nature (or the lack thereof) [see (4.4) and (4.10)] of the white noise process in TM which is above the loop $\phi$. The velocity process associated with the circular white noise from (4.10) is proved to be nothing but the usual periodic Ornstein-Uhlenbeck process-see Section 4. General description in terms of the covariance function of the class of all real two-point Markov processes indexed by the circle is given in [4]. It should be noted that, in the present context, it is essential to work with complex-valued processes and that the respective theory is very different from its real-valued counterpart if the holonomy is nontrivial.
2. Objectives and notation. Throughout, $M$ will denote a fixed Riemannian manifold of dimension $\mathrm{d} \geq 2$ endowed with the Riemannian connection and $\phi: S^{1} \mapsto \mathrm{M}$ will be some fixed loop on M . Here, $\mathrm{S}^{1}$ will be identified with the interval $[0,2 \pi$ [ in the obvious way, which, in turn, endows $\phi$ with positive and negative direction. We suppose that parallel translation along $\phi$ is possible; this is always the case if $\phi$ is smooth or is a sample path of a semimartingale. $N$ ext, let $\mathrm{V}(\theta), \theta \in \mathrm{S}^{1}, 1 \leq \mathrm{d}^{\prime}:=\operatorname{dim} \mathrm{V}(\theta) \leq \mathrm{d}$, be a continu-
ous distribution of tangent spaces along $\phi$, which is invariant under parallel translation. That is, $\forall \theta \in \mathrm{S}^{1} \mathrm{~V}(\theta)$ is a vector subspace of $\mathrm{T}_{\phi(\theta)} \mathrm{M}$ and as such is obtained by parallel translation of $\mathrm{V}(0)$ along $\phi$ in the positive direction from $\phi(0)$ to $\phi(\theta)$. The continuity assumption implies that the parallel translation of $\mathrm{V}(0)$ along the entire loop results in a rotation, possibly the trivial one, inside $\mathrm{V}(0)$. This rotation splits $\mathrm{V}(0)$ into orthogonal subspaces of dimension less than or equal to 2, every one of which is invariant under its action. Thus $\theta \leadsto \mathrm{V}(\theta)$ is just an orthogonal sum of continuous translation invariant distributions of dimension less than or equal to 2.

Our goal is to study continuous random processes $\mathrm{X}_{\theta}, \theta \in \mathrm{S}^{1}$, which take values in TM above $\phi$ and which are supported by the distribution $V$, in that $\forall \theta \in \mathrm{S}^{1} \mathrm{X}_{\theta}$ is randomly distributed in the vector space $\mathrm{V}(\theta) \subseteq \mathrm{T}_{\phi(\theta)} \mathrm{M}$. Because of the remark just made, the interesting case is $\mathrm{d}^{\prime} \leq 2$. If $\mathrm{d}^{\prime}=2$, after choosing some orthonormal frame $\varphi_{0}$ in $\mathrm{V}(0), \mathrm{X}_{\theta}, \theta \in \mathrm{S}^{1}$, could be treated as a complex-valued process $\mathrm{x}_{\theta} \in \mathbf{C}, \theta \in\left[0,2 \pi\left[\right.\right.$, chosen so that $\mathfrak{i} \mathrm{x}_{\theta}$ and $\mathfrak{\Im} x_{\theta}$ are nothing but the coördinates of $X_{\theta}$ in the frame $\varphi_{\theta}:=$ the translate of $\varphi_{0}$ along $\phi$ in positive direction from $\phi(0)$ to $\phi(\theta)$. Furthermore, $\mathrm{x}_{\theta}$ could be extended to a continuous process $\mathrm{x}_{\theta} \in \mathbf{C}, \theta \in \mathbf{R}$, with the property $\mathrm{x}(\theta+2 \pi)=\mathrm{e}^{\mathrm{i} \alpha} \mathrm{x}_{\theta} \forall \theta \in \mathbf{R}$, where $\alpha \in[0,2 \pi$ [ describes the rotation of $\varphi_{0}$ inside $\mathrm{V}(0)$ caused by parallel translation in positive direction along the entire loop $\phi$. It is to be noted that since the rotation group $\mathrm{O}(2)$ is commutative, in general, $\alpha$ will depend on the loop $\phi$ and on the tangent subspace $\mathrm{V}(0)$, but not on the frame $\varphi_{0}$. Of course, the process $x$ depends on $\varphi_{0}$ and saying that its distribution is independent of $\varphi_{0}$ is equivalent to the claim that the distribution of $X$ is invariant under the action of $O(2)$. The case $\mathrm{d}^{\prime}=1$ is easy to reduce to the one just described by taking either $\alpha=\pi$ in case the parallel translation along $\phi$ reverses the direction in $\mathrm{V}(0)$, or $\alpha=0$ if it preserves it, after which $\mathrm{X}_{\theta}, \theta \in \mathbf{R}$, could be identified with the real part of $\mathrm{X}_{\theta}, \theta \in \mathbf{R}$.

From now on, our only concern will be the case where $\mathrm{d}^{\prime}=2$ and where $\mathrm{X}_{\theta}, \theta \in \mathrm{S}^{1}$, is continuous, stationary, symmetric and Gaussian in the sense that $\mathrm{x}_{\theta}, \theta \in \mathbf{R}$, is a L2-continuous, stationary (wide sense), complex Gaussian process with $\mathbf{E}\left\{\mathrm{x}_{\theta}\right\}=\mathbf{E}\left\{\mathrm{x}_{\theta} \mathrm{x}_{\theta^{\prime}}\right\}=0, \forall \theta, \theta^{\prime} \in \mathbf{R}$. Clearly, if $\operatorname{dim} V>2$, a stationary symmetric Gaussian process supported by V is simply the sum of (appropriately normalized) such processes supported by the orthogonal distributions (of dimension less than or equal to 2); V is being split into by the hol onomy along $\phi$. So, we will be working only with the case

$$
\begin{equation*}
\mathbf{x}_{\theta}=\sum_{k \in \mathbf{Z}} \exp \left(\mathrm{i}\left(\frac{\alpha}{2 \pi}+\mathrm{k}\right) \theta\right) \xi_{\mathrm{k}}, \tag{2.1}
\end{equation*}
$$

where $\xi_{k}, \mathrm{k} \in \mathbf{Z}$, are independent complex Gaussian r.v.'s with

$$
\sum_{k \in \mathbf{Z}} \mathbf{E}\left\{\xi_{\mathrm{k}} \overline{\xi_{\mathrm{k}}}\right\}<\infty \quad \text { and } \quad \mathbf{E}\left\{\xi_{\mathrm{k}}\right\}=\mathbf{E}\left\{\xi_{\mathrm{k}}^{2}\right\}=0 \quad \forall \mathrm{k} \in \mathbf{Z}
$$

If, in addition,

$$
\sum_{k \in \mathbf{Z}}\left|\frac{\alpha}{2 \pi}+k\right|^{2} \mathbf{E}\left\{\xi_{k} \overline{\xi_{k}}\right\}<\infty,
$$

then the $L^{2}$-derivative of $\theta \leadsto \mathrm{X}_{\theta}$ (note that this is nothing but the covariant derivative of $\theta \leadsto \mathrm{X}_{\theta}$ ) is well defined and given by

$$
\begin{equation*}
\sum_{\mathrm{k} \in \mathbf{Z}} \mathrm{i}\left(\frac{\alpha}{2 \pi}+\mathrm{k}\right) \exp \left(\mathrm{i}\left(\frac{\alpha}{2 \pi}+\mathrm{k}\right) \theta\right) \xi_{\mathrm{k}} . \tag{2.2}
\end{equation*}
$$

Thus, the spectrum of $\theta \leadsto \mathrm{x}_{\theta}$ could only be inside the set $\alpha /(2 \pi)+\mathrm{k}, \mathrm{k} \in \mathbf{Z}$. Consequently, if $\xi_{\mathrm{k}}, \mathrm{k} \in \mathbf{Z}$, are chosen so that

$$
\begin{equation*}
\mathbf{E}\left\{\xi_{\mathrm{k}} \overline{\xi_{\mathrm{k}}}\right\}=\text { constant } \times\left|\frac{\alpha}{2 \pi}+\mathrm{k}\right|^{-2}, \quad \mathrm{k} \in \mathbf{Z}, \mathrm{k} \neq-\frac{\alpha}{2 \pi}, \tag{2.3}
\end{equation*}
$$

then $\theta \leadsto \nabla_{\theta} X_{\theta}(\nabla:=$ covariant derivative along $\phi)$ will have the meaning of white noise process, for in that case the spectral measure of $\theta \leadsto \mathrm{X}_{\theta}$ will be uniformly distributed on the biggest possible set the spectrum could live on. Of course, in this later case the expression in (2.2) cannot give meaning to $\theta \leadsto \nabla_{\theta} X_{\theta}$, or, equivalently, to $\theta \leadsto x(\theta)$.

To summarize, we will only consider the case where $\mathrm{d}^{\prime}=\operatorname{dim}(\mathrm{V})=2$ and where $\mathrm{x}_{\theta}, \theta \in \mathbf{R}$, is given by (2.1), assuming that (2.3) holds with constant $=1$.

The covariance of $\mathrm{x}_{\theta}, \theta \in \mathbf{R}$, is then given by

$$
\mathrm{R}\left(\theta^{\prime}-\theta\right)=\exp \left(\mathrm{i} \frac{\alpha}{2 \pi}\left(\theta^{\prime}-\theta\right)\right) \sum_{\mathrm{k} \in \mathbf{Z}} \frac{\exp \left(\mathrm{ik}\left(\theta^{\prime}-\theta\right)\right)}{|\alpha / 2 \pi+\mathrm{k}|^{2}}
$$

where $\alpha \in[0,2 \pi[$ is uniquely determined by the loop $\phi$ and the distribution $\theta \leadsto \mathrm{V}(\theta)$; if $\mathrm{V}(0)$ is identified with $\mathbf{C}$ via some orthonormal frame, then after parallel translation along the entire loop $\phi$ in positive direction any $z \in V(0)$ $\equiv \mathbf{C}$ arrives at $\mathrm{e}^{-\mathrm{i} \alpha} \mathrm{z} \in \mathrm{V}(0) \equiv \mathbf{C}$. Our main objective is to find a more tractable expression for $R(\cdot)$ and derive the Markovian nature of the process $X$, or, equivalently, of $x$.
3. The Hilbert space of $\mathbf{X}$. The case $\alpha=0$ differs in an essential way from $\alpha \neq 0$ and requires different treatment: if $\alpha \neq 0$, the path-space of x contains no constants and if $\alpha=0$, it does. First we study the simpler case $\alpha \neq 0$.

Definition 3.1. A mapping $\mathbf{R} \ni \theta \leadsto \mathrm{z}(\theta) \in \mathbf{C}$ with the property $\mathrm{z}(\theta+$ $2 \pi)=\mathrm{e}^{\mathrm{i} \alpha} \mathrm{Z}(\theta) \forall \theta \in \mathbf{R}$ will be called $\alpha$-periodic. The space of all absolutely continuous, that is, continuous and a.e. differentiable, $\alpha$-periodic complex-valued functions on $\mathbf{R}$ will be denoted by $\mathrm{H}_{\alpha}$ and will be endowed with the usual scalar product

$$
\langle\mathrm{h} \mid \mathrm{k}\rangle:=\frac{1}{2 \pi} \int_{[0,2 \pi \mathrm{~L}} \mathrm{h}^{\prime}(\theta) \overline{\mathrm{k}^{\prime}(\theta)} \mathrm{d} \theta, \quad \mathrm{~h}, \mathrm{k} \in \mathrm{H}_{\alpha} .
$$

In fact, $\mathrm{H}_{\alpha}$ consists of absolutely continuous mappings from $\mathrm{S}^{1}$ into V as explained in the following remark.

Remark 3.2. Let $\varphi_{0} \equiv\left(\mathrm{e}_{0}^{\mathrm{R}}, \mathrm{e}_{0}^{\mathrm{l}}\right)$, be some orthonormal frame in $\mathrm{V}(0)$ and let $\varphi_{\theta} \equiv\left(\mathrm{e}_{\theta}^{\mathrm{R}}, \mathrm{e}_{\theta}^{\mathrm{l}}\right):=$ the parallel transport of $\varphi_{0}$ along $\phi$ from $\mathrm{T}_{\phi(0)} \mathrm{M}$ to $\mathrm{T}_{\phi(\theta)} \mathrm{M}$. Given $z \in \mathbf{C}$, set $\varphi_{\theta}[z]:=(\Re z) \mathrm{e}_{\theta}^{\mathrm{R}}+(\Im z) \mathrm{e}_{\theta}^{\prime} \in \mathrm{V}(\theta)$ and notice that if $\mathbf{R} \ni \theta$ $\leadsto \mathbf{z}(\theta) \in \mathbf{C}$ is $\alpha$-periodic and continuous then

$$
\left[0,2 \pi\left[\ni \theta \leadsto(\varphi \mathrm{z})(\theta):=\varphi_{\theta}[\mathrm{z}(\theta)] \in \mathrm{V}(\theta)\right.\right.
$$

is continuous too. Thus, the correspondence $\mathrm{H}_{\alpha} \ni \mathrm{Z}(\varphi \mathrm{z}) \in \mathrm{V}$ allows identifying $H_{\alpha}$ with the space of all continuous vector fields above $\phi$ which are supported by V and admit covariant derivatives at almost every point on the loop $\phi$. This association will be often blurred in the sequel, and, with a slight abuse of terminology and notation, continuous complex $\alpha$-periodic functions on $\mathbf{R}$ will be treated also as functions from $\mathrm{S}^{1}$ into TM supported by the distribution $\theta \leadsto \mathrm{V}(\theta)$. The space of all such functions will be denoted by $\mathrm{C}_{\alpha}\left(\mathrm{S}^{1} \rightarrow \mathrm{~V}\right)$ and we will no longer distinguish between the $\mathbf{C}$-valued process x and the process $\mathrm{X}=\varphi \mathrm{X} \in \mathrm{V}$.

For $\mathrm{t} \in\left[0,2 \pi\left[\right.\right.$ let $\mathrm{T}_{\mathrm{t}}: \mathrm{H}_{\alpha} \mapsto \mathrm{H}_{\alpha}$ be the usual shift operator $\mathrm{T}_{\mathrm{t}} \mathrm{h}(\theta):=\mathrm{h}(\theta-\mathrm{t})$, $\theta \in \mathbf{R}$, and given $\mathrm{h} \in \mathrm{H}_{\alpha}$, consider the vector field

$$
\rho_{\mathrm{t}}(\varphi \mathrm{~h})(\theta):=\varphi_{\theta}\left[\mathrm{T}_{\mathrm{t}} \mathrm{~h}(\theta)\right] \in \mathrm{V}(\theta), \quad \theta \in[0,2 \pi[.
$$

It is easy to see that $\rho_{\mathrm{t}}(\varphi \mathrm{h})(\cdot)$ is obtained by rotating (by way of parallel transport) the vector field $(\varphi h)(\cdot)$ along $\phi$ in positive direction; for example $\rho_{\mathrm{t}}(\varphi \mathrm{h})(\mathrm{t})$ is obtained by parallel translation of $(\varphi \mathrm{h})(0)$ from $\mathrm{T}_{\phi(0)} \mathrm{M}$ to $\mathrm{T}_{\phi(\mathrm{t})} \mathrm{M}$.

Next, given $t \in[0,2 \pi[$, let

$$
\Lambda_{\mathrm{t}}(\theta):=\left[\mathrm{e}^{\mathrm{i} \alpha} 1_{[0, \mathrm{t}}(\theta)+1_{[\mathrm{t}, 2 \pi[ }(\theta)\right] \frac{\mathrm{t}-\theta}{1-\mathrm{e}^{\mathrm{i} \alpha}}-\frac{2 \pi \mathrm{e}^{\mathrm{i} \alpha}}{\left(1-\mathrm{e}^{\mathrm{i} \alpha}\right)^{2}}, \quad \theta \in[0,2 \pi[
$$

[we assume that for $\mathrm{t}=0\left[0, \mathrm{t}\left[=\varnothing\right.\right.$ and $\left.1_{[0, \mathrm{t}[ }(\theta) \equiv 0\right]$, extend $\Lambda_{\mathrm{t}}$ to an $\alpha$-periodic complex function on $\mathbf{R}$ and notice that, so defined, $\Lambda_{\mathrm{t}}$ is in fact an element of $H_{\alpha}$ and that $\Lambda_{\mathrm{t}}=\mathrm{T}_{\mathrm{t}} \Lambda_{0} \forall \mathrm{t} \in\left[0,2 \pi\left[\right.\right.$; equivalently, $\varphi \Lambda_{\mathrm{t}}=\rho_{\mathrm{t}}\left(\varphi \Lambda_{0}\right)$, for any choice of frame $\varphi$ in $\mathrm{V}(0) \subset \mathrm{T}_{\phi(0)} \mathrm{M}$. More importantly, notice that

$$
\left[0,2 \pi\left[\backslash\{\mathrm{t}\} \ni \theta \rightsquigarrow-\Lambda_{\mathrm{t}}^{\prime}(\theta)=\frac{\mathrm{e}^{\mathrm{i} \alpha}}{1-\mathrm{e}^{\mathrm{i} \alpha}} \mathbf{1}_{[0, \mathrm{tI}}(\theta)+\frac{1}{1-\mathrm{e}^{\mathrm{i} \alpha}} \mathbf{1}_{[\mathrm{t}, 2 \pi[ }(\theta) \in \mathbf{C}\right.\right.
$$

is just the circular Heaviside function with jump at $t \in[0,2 \pi[$, in that $-\Lambda_{\mathrm{t}}^{\prime}(\theta)$ is constant on $\mathrm{S}^{1} \backslash \mathrm{t}$ and

$$
\lim _{\theta>\mathrm{t}}-\Lambda_{\mathrm{t}}^{\prime}(\theta)+\lim _{\theta \nearrow \mathrm{t}} \Lambda_{\mathrm{t}}^{\prime}(\theta)=1 ;
$$

consequently, $\Lambda_{\mathrm{t}}(\cdot)$ is an integral kernel for the evaluation map $H_{\alpha} \ni \mathrm{h} \leadsto$ $h(\mathrm{t}) \in \mathbf{C}$ in the sense of the following proposition.

Proposition 3.3. For any choice of $h \in H_{\alpha}$ and $t \in[0,2 \pi[$, one has

$$
\left\langle\mathrm{h} \mid 2 \pi \Lambda_{\mathrm{t}}\right\rangle \equiv \int_{[0,2 \pi \mathrm{~h}} \mathrm{h}^{\prime}(\theta) \overline{\Lambda_{\mathrm{t}}^{\prime}(\theta)} \mathrm{d} \theta=\mathrm{h}(\mathrm{t}) \in \mathbf{C} .
$$

Next, notice that

$$
\lambda_{\mathrm{k}}(\theta):=\frac{\exp (\mathrm{i}(\alpha / 2 \pi+\mathrm{k}) \theta)}{\mathrm{i}(\alpha / 2 \pi+\mathrm{k})}, \quad \mathrm{k} \in \mathbf{Z},
$$

form an orthonormal basis in $H_{\alpha}$, consisting, in fact, of eigenfunctions of the Laplacian $\Delta=\mathrm{d}^{2} / \mathrm{d} \theta^{2}$ and that, by choosing $\xi_{\mathrm{k}}, \mathrm{k} \in \mathbf{Z}$, as in (2.3) with constant $=1$, the process $x_{t}, t \in[0,2 \pi[$, given by (2.1) could be written as

$$
\begin{equation*}
\mathrm{X}_{\mathrm{t}}=\sum_{\mathrm{k} \in \mathbf{Z}} \frac{\exp (\mathrm{i}(\alpha / 2 \pi+\mathrm{k}) \mathrm{t})}{\mathrm{i}(\alpha / 2 \pi+\mathrm{k})} \gamma_{\mathrm{k}} \equiv \sum_{\mathrm{k} \in \mathbf{Z}}\left\langle\lambda_{\mathrm{k}} \mid 2 \pi \Lambda_{\mathrm{t}}\right\rangle \gamma_{\mathrm{k}}, \tag{3.1}
\end{equation*}
$$

where $\gamma_{\mathrm{k}}, \mathrm{k} \in \mathbf{Z}$, are i.i.d. complex Gaussian r.v.'s with $\mathbf{E}\left\{\gamma_{\mathrm{k}}\right\}=\mathbf{E}\left\{\gamma_{\mathrm{k}}^{2}\right\}=0$ and $\mathbf{E}\left\{\gamma_{k} \overline{\gamma_{k}}\right\}=1$. Plainly, if $\mathrm{C}_{\alpha}\left(\mathrm{S}^{1} \rightarrow \mathrm{~V}\right)$ is being treated as a probability space equipped with the Borel $\sigma$-field and the Gaussian probability law $\wp$ derived from the canonical inclusion $\mathrm{H}_{\alpha} \rightarrow \mathrm{C}_{\alpha}\left(\mathrm{S}^{1} \rightarrow \mathrm{~V}\right)$ and the associated abstract Wiener space construction, then the process $\mathrm{X}_{\theta}, \theta \in \mathbf{R}$, and the complex-valued random process

$$
\mathbf{R} \ni \theta \leadsto 2 \pi z(\theta) \in \mathbf{C}, \quad z \in \mathcal{C}_{\alpha}\left(S^{1} \rightarrow V\right),
$$

are indistinguishable in the sense that their respective finite-dimensional distributions are identical. In particular, this shows that

$$
\begin{align*}
\mathrm{R}(\mathrm{t}) & :=\frac{1}{2 \pi} \mathbf{E}\left\{\mathrm{X}_{0} \overline{\mathrm{X}}_{\mathrm{t}}\right\}=\left\langle\Lambda_{0} \mid 2 \pi \Lambda_{\mathrm{t}}\right\rangle=\Lambda_{0}(\mathrm{t}) \\
& \equiv \frac{\mathrm{t}}{\mathrm{e}^{\mathrm{i} \alpha}-1}+\frac{\pi}{1-\cos \alpha}, \quad \mathrm{t} \in[0,2 \pi[. \tag{3.2}
\end{align*}
$$

Now we develop the counterpart of the above for the case $\alpha=0$. In view of Definition 3.1, $\mathrm{H}:=\mathrm{H}_{\alpha=0}$ is simply the space of absolutely continuous and periodic (with period $2 \pi$ ) complex functions on $\mathbf{R}$; that is, the space of absolutely continuous functions on the circle. $\langle\cdot \mid \cdot\rangle^{1 / 2}$ is no longer a norm and we turn H into a Hilbert space by decomposing it into the sum

$$
H=\mathbf{C}+K, \quad K:=\left\{h \in H \mid \int_{[0,2 \pi[ } h=0\right\},
$$

and by setting

$$
\left(\mathrm{c}_{1}+\mathrm{h}_{1} \mid \mathrm{c}_{2}+\mathrm{h}_{2}\right):=\mathrm{c}_{1} \overline{\mathrm{c}_{2}}+\left\langle\mathrm{h}_{1} \mid \mathrm{h}_{2}\right\rangle \equiv \mathrm{c}_{1} \overline{\mathrm{c}_{2}}+\frac{1}{2 \pi} \int_{[0,2 \pi[ } \mathrm{h}_{1}^{\prime}(\theta) \overline{\mathrm{h}_{2}^{\prime}(\theta)} \mathrm{d} \theta,
$$

$c_{1}, c_{2} \in \mathbf{C}, h_{1}, h_{2} \in K$. Note that $\langle\cdot \mid \cdot\rangle^{1 / 2}$ is a norm on $K$ and therefore $H$ turns into a Hilbert space with norm $(\cdot \mid \cdot)^{1 / 2}$. Similarly, by splitting each $z \in \mathrm{C}_{\alpha=0}\left(\mathrm{~S}^{1} \rightarrow \mathrm{~V}\right)$ into the sum

$$
z(\theta)=\frac{1}{2 \pi} \int_{[0,2 \pi[ } z(\theta) d \theta+\left(z(\theta)-\frac{1}{2 \pi} \int_{[0,2 \pi[ } z(\theta) d \theta\right), \quad \theta \in[0,2 \pi[,
$$

$\mathrm{C}_{0}\left(\mathrm{~S}^{1} \rightarrow \mathrm{~V}\right)$ could be written as $\mathbf{C}+\tilde{\mathrm{C}}_{0}\left(\mathrm{~S}^{1} \rightarrow \mathrm{~V}\right)$, the second component of which is endowed with the standard sup-norm. We suppose that $\mathrm{C}_{0}\left(\mathrm{~S}^{1} \rightarrow \mathrm{~V}\right)$ has the topol ogy of the product $\mathbf{C} \otimes \tilde{C}_{0}\left(\mathrm{~S}^{1} \rightarrow \mathrm{~V}\right)$ and endow $\mathrm{C}_{0}\left(\mathrm{~S}^{1} \rightarrow \mathrm{~V}\right)$ with probability law $\wp$ which is the product of two independent probability laws $\Gamma$ and $\tilde{\wp}$, respectively, on $\mathbf{C}$ and on $\tilde{C}_{0}\left(\mathrm{~S}^{1} \rightarrow \mathrm{~V}\right), \Gamma:=$ the law of a complex Gaussian r.v. $\zeta$ with $\mathbf{E}\{\zeta\}=\mathbf{E}\left\{\zeta^{2}\right\}=0$ and $\tilde{\wp}:=$ the Gaussian measure on $C_{0}\left(\mathrm{~S}^{1} \rightarrow \mathrm{~V}\right)$ derived from the abstract Wiener space associated with the inclusion $\mathrm{K} \rightarrow \mathrm{C}_{0}\left(\mathrm{~S}^{1} \rightarrow \mathrm{~V}\right)$. To this end, we consider the space K as a Hilbert space with scalar product $\langle\cdot \mid \cdot\rangle$ and choose an orthonormal basis in K formed by the eigenfunctions of the Laplacian (Laplacian on $K$, that is),

$$
\lambda_{\mathrm{k}}(\theta):=\frac{1}{\mathrm{ik}} \mathrm{e}^{\mathrm{ik} \theta}, \quad \mathrm{k} \in \mathbf{Z} \backslash\{0\} .
$$

In the present setting the circular Heavisidefunction with jump at $\mathrm{t} \in[0,2 \pi[$ is ( -1 ) times the derivative of the following quadratic (rather than linear, as was the case where $\alpha \neq 0$ ) function:

$$
\Lambda_{\mathrm{t}}(\theta):=\frac{1}{4 \pi}(\theta-\mathrm{t})^{2}-\frac{1}{2}(\theta-\mathrm{t})+(\theta-\mathrm{t}) 1_{[0, \mathrm{t}[ }(\theta)+\frac{\pi}{6}, \quad \theta \in[0,2 \pi[
$$

which, just as before, we extend to a periodic function on $\mathbf{R}$ and remark that so defined it is an element of K and that $\Lambda_{\mathrm{t}}(\theta)=\mathrm{T}_{\mathrm{t}} \Lambda_{0}(\theta)=\Lambda_{0}(\theta-\mathrm{t})$. Obviously, $\Lambda_{\mathrm{t}}^{\prime}$ is no longer a constant on $[0,2 \pi[\backslash\{\mathrm{t}\}$ and the term "Heaviside function" is justified by

Proposition 3.4. For any choice of $h \in K$ and $t \in[0,2 \pi[$ one has

$$
\left\langle\mathrm{h} \mid 2 \pi \Lambda_{\mathrm{t}}\right\rangle \equiv \int_{[0,2 \pi \mathrm{~L}} \mathrm{h}^{\prime}(\theta) \overline{\Lambda_{\mathrm{t}}^{\prime}(\theta)} \mathrm{d} \theta=\mathrm{h}(\mathrm{t}) \in \mathbf{C}
$$

where

$$
\Lambda_{\mathrm{t}}^{\prime}(\theta)=\frac{1}{2 \pi}(\theta-\mathrm{t})-\frac{1}{2}+1_{\left[0, \mathrm{t}_{[ }\right.}(\theta), \quad \theta \in[0,2 \pi[\backslash\{\mathrm{t}\} .
$$

Proof. It is enough to notice that

$$
\Lambda_{\mathrm{t}}^{\prime \prime}(\theta)=-\delta_{\mathrm{t}}(\theta)+(1 / 2 \pi) 1_{[0,2 \pi \backslash \backslash\{\mathrm{t}]}(\theta), \quad \theta \in[0,2 \pi[.
$$

Consequently, (3.1) now becomes

$$
\begin{align*}
X_{\mathrm{t}} & =\xi_{0}+\sum_{\mathrm{k} \in \mathbf{Z} \backslash\{0\}} \frac{1}{\mathrm{ik}} \mathrm{e}^{\mathrm{ikt}} \gamma_{\mathrm{k}}  \tag{3.3}\\
& \equiv \xi_{0}+\sum_{\mathrm{k} \in \mathbf{Z} \backslash\{0\}}\left\langle\lambda_{\mathrm{k}} \mid 2 \pi \Lambda_{\mathrm{t}}\right\rangle \lambda_{\mathrm{k}}=: \xi_{0}+\mathrm{Y}_{\mathrm{t}}
\end{align*}
$$

and (3.2) turns into

$$
\begin{align*}
\mathrm{R}(\mathrm{t}) & :=\frac{1}{2 \pi} \mathbf{E}\left\{\mathrm{Y}_{0} \overline{\mathrm{Y}}_{\mathrm{t}}\right\}=\left\langle\Lambda_{0} \mid 2 \pi \Lambda_{\mathrm{t}}\right\rangle=\Lambda_{0}(\mathrm{t}) \\
& \equiv \frac{\mathrm{t}^{2}}{4 \pi}-\frac{\mathrm{t}}{2}+\frac{\pi}{6}, \quad \mathrm{t} \in[0,2 \pi[. \tag{3.4}
\end{align*}
$$

3.5 Remark. 1. The Hilbert space associated with the process $Y_{t}:=X_{t}-$ $\xi_{0}, \mathrm{t} \in \mathrm{S}^{1}$, is K , not $H$.
2. Since the covariance function in (3.2) is real for $\alpha=\pi$, and so is the one in (3.4), it is easy to conclude from the above that if $X_{t}^{\prime}, t \in S^{1}$, is a real-valued (Gaussian) white noise process then, up to a factor, the covariance of $X_{t}, t \in S^{1}$, is either

$$
\mathrm{R}(\mathrm{t})=\mathrm{E}\left\{\mathrm{X}_{0} \mathrm{X}_{\mathrm{t}}\right\}=-\frac{\mathrm{t}}{2}+\frac{\pi}{2}, \quad \mathrm{t} \in[0,2 \pi[
$$

[this is (3.2) with $\alpha=\pi$ ] if X is antiperiodic $\left(\mathrm{X}_{\mathrm{t}+2 \pi}=-\mathrm{X}_{\mathrm{t}}\right)$, or is given by

$$
\frac{\mathrm{t}^{2}}{4 \pi}-\frac{\mathrm{t}}{2}+\frac{\pi}{6}+\text { constant, } \mathrm{t} \in[0,2 \pi[
$$

with some arbitrary real positive constant, if $X$ is periodic $\left(X_{t+2 \pi}=X_{t}\right)$. Note that in this later case the covariance is not unique (unique up to a factor, that is).
4. The Markovian nature of $\mathbf{X}$. Here again we treat the case $\alpha \neq 0$ and $\alpha=0$ separately and consider the case $\alpha \neq 0$ first.

As scaling makes no difference in our study, we assume that $X_{t}, t \in[0,2 \pi[$, has covariance $R(t)$, $t \in[0,2 \pi[$, given by (3.2). For $t<0$ we set $R(t)$ $:=\overline{\mathrm{R}(-\mathrm{t})}$ and notice that

$$
\begin{equation*}
\frac{\partial^{2}}{\partial \mathrm{t} \partial \mathrm{~s}} \mathrm{R}(\mathrm{t}-\mathrm{s})=\delta_{0}(\mathrm{t}-\mathrm{s}), \quad \mathrm{s}, \mathrm{t} \in[0,2 \pi[, \tag{4.1}
\end{equation*}
$$

and therefore that stochastic integrals of the form

$$
\int_{[0,2 \pi[ } f(t) d X_{t}, \quad f \in L^{2}\left(S^{1} \rightarrow \mathbf{C}\right)
$$

are well defined as is the quadratic variation $d X_{t} \overline{\mathrm{dX}}=\mathrm{dt}$. Given $0 \leq \mathrm{s}<\mathrm{t}<$ $2 \pi$, we have

$$
E\left\{X_{0} \mid X_{s}\right\}=\frac{R(s)}{R(0)} X_{s} \text { and } E\left\{X_{t} \mid X_{s}\right\}=\frac{\overline{R(t-s)}}{R(0)} X_{s}
$$

and so

$$
E\left\{X_{t} \mid X_{0}, X_{s}\right\}=\varkappa(s, t)\left(X_{0}-\frac{R(s)}{R(0)} X_{s}\right)+\frac{\overline{R(t-s)}}{R(0)} X_{s}
$$

where $x(\mathrm{~s}, \mathrm{t})$ is to be determined by the condition

$$
E\left\{\left(X_{t}-E\left\{X_{t} \mid X_{0}, X_{s}\right\}\right) \bar{X}_{0}\right\}=0
$$

Somewhat tedious but otherwise trivial calculation yields

$$
x(s, t)=\frac{\overline{R(t)} R(0)-\overline{R(t-s) R(s)}}{R(0)^{2}-|R(s)|^{2}}=e^{i \alpha} \frac{t-s}{2 \pi-s}
$$

from which one finds that

$$
\begin{equation*}
E\left\{X_{t}-X_{s} \mid X_{0}, X_{s}\right\}=\frac{e^{i \alpha} X_{0}-X_{s}}{2 \pi-s}(t-s) \tag{4.2}
\end{equation*}
$$

Now it is easy to see that $X$ is a reciprocal process in the sense of the following proposition.

Proposition 4.1. Given $0 \leq \mathrm{s}<\mathrm{t}<2 \pi$, one has

$$
E\left\{X_{t} \mid \mathrm{X}_{\mathrm{u}}, \mathrm{u} \in[0, \mathrm{~s}]\right\}=\mathbf{E}\left\{\mathrm{X}_{\mathrm{t}} \mid \mathrm{X}_{0}, \mathrm{X}_{\mathrm{s}}\right\} .
$$

Proof. The claim is that

$$
\mathbf{E}\left\{\left(X_{\mathrm{t}}-\mathbf{E}\left\{\mathrm{X}_{\mathrm{t}} \mid X_{0}, X_{\mathrm{s}}\right\}\right) \overline{X_{\mathrm{u}}}\right\}=0 \quad \forall \mathrm{u} \in[0, \mathrm{~s}]
$$

or, what amounts to the same thing, that

$$
\begin{gather*}
\overline{R(t-u)} R(0)-x(s, t)[R(u) R(0)-R(s) \overline{R(s-u)}]  \tag{4.3}\\
-\overline{R(t-s) R(s-u)}=0 \quad \forall u \in[0, s]
\end{gather*}
$$

which is certainly true, because the left-hand side is linear in u and $\varkappa(\mathrm{s}, \mathrm{t})$ was chosen so that the identity holds for $\mathrm{u}=0$ and $\mathrm{u}=\mathrm{s}$.

Remark 4.2. For a wide sensestationary, $\mathrm{L}^{2}$-continuous, periodic Gaussian process with period $2 \pi$ and covariance $R$, s.t., $R(t) \neq R(0), \forall t \in[0,2 \pi[$, (4.3) is just a rephrasing of the two-point Markov property. It is obvious, then, that the two-point Markov property always holds if R is linear. It is easy to see that $R^{\prime \prime}(t)=$ constant $\times R(t), t \neq 0$, implies two-point Markov too, for, in that case, as a function of $u \in[0, s]$, the left side of (4.3) could only be of the form $\mathrm{Ae}^{-\mathrm{Cu}}+\mathrm{Be}^{+\mathrm{Cu}}$ and could vanish for $\mathrm{u}=0$ and for $\mathrm{u}=\mathrm{s} \neq 0$ only if $\mathrm{A}=\mathrm{B}=0$.

In fact, Proposition 4.1, combined with (4.2), comes down to the following statement.

Proposition 4.3. A continuous, stationary, complex-Gaussian stochastic process $X_{t}, t \in\left[0,2 \pi\left[\right.\right.$, with $E\left\{X_{t}\right\}=E\left\{X_{t} X_{t^{\prime}}\right\}=0$ and with covariance function $R(t):=E\left\{X_{0} X_{t}\right\}$ given by the right side of (2.2) solves the equation

$$
\left\{\begin{array}{l}
\mathrm{dX}_{\mathrm{t}}=\frac{\mathrm{e}^{\mathrm{i} \alpha} \mathrm{X}_{0}-\mathrm{X}_{\mathrm{t}}}{2 \pi-\mathrm{t}} \mathrm{dt}+\mathrm{dW}_{\mathrm{t}}, \quad 0 \leq \mathrm{t}<2 \pi,  \tag{4.4}\\
\mathrm{X}_{0}=\text { complex Gaussian r.v. with } \mathbf{E}\left\{\mathrm{X}_{0}\right\}=\mathbf{E}\left\{\mathrm{X}_{0}^{2}\right\}=0 \\
\quad \text { and } \mathbf{E}\left\{\mathrm{X}_{0} \bar{X}_{0}\right\}=\frac{\pi}{1-\cos \alpha},
\end{array}\right.
$$

driven by complex Brownian motion $W$ of intensity $d W_{t} \overline{d W}_{t}=d t$, which is independent of the initial value $X_{0}$.

This, of course, is pinned Brownian motion with appropriately randomized initial value. This leads to the following conclusion.

Corollary 4.4. The solution of (4.4) could be written in the (W-nonanticipative) form

$$
\begin{equation*}
\mathrm{X}_{\mathrm{t}} \equiv(2 \pi-\mathrm{t}) \int_{0}^{\mathrm{t}} \frac{\mathrm{dW}}{\mathrm{~s}} \text { } \frac{2 \pi+\mathrm{s}}{2 \pi-\frac{\left.\mathrm{e}^{\mathrm{i} \alpha}-1\right) \mathrm{t}}{2 \pi} \mathrm{X}_{0}, \quad \mathrm{t} \in[0,2 \pi[, ~, ~, ~} \tag{4.5a}
\end{equation*}
$$

or, equivalently, in the (W-anticipative) form

$$
\begin{equation*}
\mathrm{X}_{\mathrm{t}}=\frac{1}{\mathrm{e}^{\mathrm{i} \alpha}-1} \mathrm{~W}_{2 \pi}+\mathrm{W}_{\mathrm{t}}, \quad \mathrm{t} \in[0,2 \pi[. \tag{4.5b}
\end{equation*}
$$

It is not hard to compute $\mathrm{R}(\mathrm{t}):=\mathbf{E}\left\{\mathrm{X}_{0} \mathrm{X}_{\mathrm{t}}\right\}$ directly from this last representation and, of course, arrive at the expression in (3.2).

Now we work out the case $\alpha=0$. Instead of studying the process $X_{\mathrm{t}}$, $t \in \mathbf{R}$, we study $\mathrm{Y}_{\mathrm{t}}:=\mathrm{X}_{\mathrm{t}}-\xi_{0}$ and assume that its covariance is given by the right side of (3.4). The above calculation with $\mathrm{R}(\mathrm{t})=\mathrm{t}^{2} / 4 \pi-\mathrm{t} / 2+\pi / 6$, $t \in[0,2 \pi[$, implies that (4.3) holds with

$$
\begin{equation*}
\varkappa(\mathrm{s}, \mathrm{t})=\frac{(\pi / 6) \mathrm{R}(\mathrm{t})-\mathrm{R}(\mathrm{t}-\mathrm{s}) \mathrm{R}(\mathrm{~s})}{\pi^{2} / 36-\mathrm{R}^{2}(\mathrm{~s})}, \quad 0 \leq \mathrm{s}<\mathrm{t}<2 \pi . \tag{4.6}
\end{equation*}
$$

The left side of (4.3) is a quadratic function of $u$ which, due to the choice of $x$, vanishes for $u=0$ and $u=s$ and therefore could be $\equiv 0 \forall u \in[0, s]$, only if its second derivative w.r.t. u,

$$
\frac{1}{12}-x(\mathrm{~s}, \mathrm{t})\left[\frac{1}{12}-\mathrm{R}(\mathrm{~s}) \frac{1}{2 \pi}\right]-\mathrm{R}(\mathrm{t}-\mathrm{s}) \frac{1}{2 \pi},
$$

vanishes in $[0, s]$. This however is easily seen to fail with the implication that $Y_{t}, t \in S^{1}$, is not a reciprocal process and that $\mathbf{E}\left\{Y_{t} \mid Y_{u}, u \in[0, s]\right\}$ could be found only in the form

$$
a Y_{0}+b Y_{s}+\int_{[0, s]} f(u) Y_{u} d u,
$$

with some $a \equiv a(s, t), b \equiv b(s, t) \in \mathbf{C}$ and $f \equiv f_{s, t} \in C([0,2 \pi[\rightarrow \mathbf{C})$. Since

$$
E\left\{\int_{[0, s]} f(u) Y_{u} d u \mid Y_{s}\right\}=\left(\frac{6}{\pi} \int_{[0, s]} f(u) R(s-u) d u\right) Y_{s}
$$

we have

$$
\begin{aligned}
E\left\{\int_{[0, s]} f(y) Y_{u} d u \mid Y_{0}, Y_{s}\right\}= & \kappa\left(Y_{0}-\frac{6}{\pi} R(s) Y_{s}\right) \\
& +\left(\frac{6}{\pi} \int_{[0, s]} f(u) R(s-u) d u\right) Y_{s},
\end{aligned}
$$

where

$$
\kappa \equiv \kappa(\mathrm{s}, \mathrm{t}):=\frac{(\pi / 6) \int_{0}^{\mathrm{s}} \mathrm{f}(\mathrm{u}) \mathrm{R}(\mathrm{u}) \mathrm{du}-\mathrm{R}(\mathrm{~s}) \int_{0}^{\mathrm{s}} \mathrm{f}(\mathrm{u}) \mathrm{R}(\mathrm{~s}-\mathrm{u}) \mathrm{du}}{\left(\pi^{2} / 36\right)-\mathrm{R}^{2}(\mathrm{~s})}
$$

is found from the condition

$$
\mathbf{E}\left\{\left[\int_{[0, s]} f(u) Y_{u} d u-E\left\{\int_{[0, s]} f(u) Y_{u} d u \mid Y_{0}, Y_{s}\right\}\right] Y_{0}\right\}=0 .
$$

Consequently,

$$
\begin{aligned}
Y_{t}-E\left\{Y_{t} \mid Y_{u}, u \in[0, s]\right\}= & Y_{t}-x\left(Y_{0}-\frac{6}{\pi} R(s) Y_{s}\right)-\frac{6}{\pi} R(t-s) Y_{s} \\
& -\int_{[0, s]} f(u) Y_{u} d u+\kappa\left(Y_{0}-\frac{6}{\pi} R(s) Y_{s}\right) \\
& +\left(\frac{6}{\pi} \int_{[0, s]} f(u) R(s-u) d u\right) Y_{s},
\end{aligned}
$$

where $x \equiv x(\mathrm{~s}, \mathrm{t})$ is given by (4.6), and this identity determines completely the function $[0, s] \ni u \leadsto f(u) \equiv f_{s, t}(u)$. Indeed,

$$
\mathbf{E}\left\{\left[\mathrm{Y}_{\mathrm{t}}-\mathbf{E}\left\{\mathrm{Y}_{\mathrm{t}} \mid \mathrm{Y}_{\mathrm{u}}, \mathrm{u} \in[0, \mathrm{~s}]\right\}\right] \mathrm{Y}_{\sigma}\right\}=0 \quad \forall \sigma \in[0, \mathrm{~s}]
$$

yields

$$
\begin{align*}
0= & \frac{\pi}{6} \mathrm{R}(\mathrm{t}-\sigma)-\varkappa\left(\frac{\pi}{6} \mathrm{R}(\sigma)-\mathrm{R}(\mathrm{~s}) \mathrm{R}(\mathrm{~s}-\sigma)\right) \\
& -\mathrm{R}(\mathrm{t}-\mathrm{s}) \mathrm{R}(\mathrm{~s}-\sigma) \\
& -\frac{\pi}{6} \int_{[0, \mathrm{~s}]} \mathrm{f}(\mathrm{u}) \mathrm{R}(\sigma-\mathrm{u}) \mathrm{du}+\kappa\left(\frac{\pi}{6} \mathrm{R}(\sigma)-\mathrm{R}(\mathrm{~s}) \mathrm{R}(\mathrm{~s}-\sigma)\right)  \tag{4.7}\\
& +\left(\int_{[0, \mathrm{~s}]} \mathrm{f}(\mathrm{u}) \mathrm{R}(\mathrm{~s}-\mathrm{u}) \mathrm{du}\right) \mathrm{R}(\mathrm{~s}-\sigma) .
\end{align*}
$$

Now, treated as functions of $\sigma$ for fixed s and t , all terms in the above expression are quadratic with the exception of $\int_{0}^{s} f(u) R(\sigma-u) d u$ and since

$$
\begin{equation*}
\mathrm{R}^{\prime \prime}(\mathrm{u})=-\delta_{0}(\mathrm{u})+\frac{1}{2 \pi} \tag{4.8}
\end{equation*}
$$

it follows that

$$
\frac{d^{2}}{d \sigma^{2}} \int_{[0, s]} f(u) R(\sigma-u) d u=-f(\sigma)+\frac{1}{2 \pi} \int_{[0, s]} f(u) d u .
$$

Thus, differentiating (4.7) w.r.t. $\sigma$ twice, one sees that actually [0, s] $\ni \mathrm{u} \leadsto$ $\mathrm{f}(\mathrm{u}) \equiv \mathrm{f}_{\mathrm{s}, \mathrm{t}}(\mathrm{u})$ is a constant. On the other hand, thanks to the choice of $\varkappa$ and $\kappa$, (4.7) is automatically satisfied for $\sigma=0$ and $\sigma=\mathrm{s}$, so that (4.7) hol ds for $\forall$ $\sigma \in[0, \mathrm{~s}]$ if and only if the second derivative w.r.t. $\sigma$ of the expression in the right side is identically null. This allows calculating the constant $f_{s, t}$ :
$f_{s, t}=\frac{(1 / 12)-(1 / 2 \pi) x[\pi / 6-R(s)]-(1 / 2 \pi) R(t-s)}{(\pi / 6)(s / 2 \pi-1)-\kappa^{\prime}(1 / 2 \pi)[\pi / 6-R(s)]-(1 / 2 \pi) \int_{[0, s]} R(u) d u}$,
where

$$
\kappa^{\prime}=\frac{\int_{[0, s \mathrm{~s}} \mathrm{R}(\mathrm{u}) \mathrm{du}}{\pi / 6+\mathrm{R}(\mathrm{~s})}
$$

and

$$
\int_{[0, s]} R(u) d u \equiv \int_{[0, s]} R(s-u) d u=\frac{s(s-\pi)(s-2 \pi)}{12 \pi} .
$$

Cumbersome as this last expression may appear, it simplifies to something surprisingly simple:

$$
f_{s, t}=\frac{6(t-s)(t-2 \pi)}{(2 \pi-s)^{3}}
$$

and one finds that

$$
\begin{aligned}
\mathbf{E}\left\{\mathrm{Y}_{\mathrm{t}}-\right. & \left.\mathrm{Y}_{\mathrm{s}} \mid \mathrm{Y}_{\mathrm{u}}, \mathrm{u} \in[0, \mathrm{~s}]\right\} \\
= & (\varkappa-\kappa)\left(\mathrm{Y}_{0}-\frac{6}{\pi} \mathrm{R}(\mathrm{~s}) \mathrm{Y}_{\mathrm{s}}\right)+\frac{6}{\pi}\left(\mathrm{R}(\mathrm{t}-\mathrm{s})-\frac{\pi}{6}\right) \mathrm{Y}_{\mathrm{s}} \\
& -\frac{6}{\pi} \frac{6(\mathrm{t}-\mathrm{s})(\mathrm{t}-2 \pi)}{(2 \pi-\mathrm{s})^{3}} \frac{\mathrm{~s}(\mathrm{~s}-\pi)(\mathrm{s}-2 \pi)}{12 \pi} \mathrm{Y}_{\mathrm{s}} \\
& +\frac{6(\mathrm{t}-\mathrm{s})(\mathrm{t}-2 \pi)}{(2 \pi-\mathrm{s})^{3}} \int_{0}^{\mathrm{s}} \mathrm{Y}_{\mathrm{u}} \mathrm{du} .
\end{aligned}
$$

It is not hard to check that

$$
\varkappa(\mathrm{s}, \mathrm{t})-\kappa(\mathrm{s}, \mathrm{t})=-\frac{2(\mathrm{t}-\mathrm{s})}{2 \pi-\mathrm{s}}+\mathrm{O}\left((\mathrm{t}-\mathrm{s})^{2}\right)
$$

and, knowing that $Y_{t}$ has quadratic variation $d Y_{t} \overline{d Y}_{t}=d t$ [see (4.8)], conclude from (4.9) that

Proposition 4.5. A continuous, stationary, complex-Gaussian stochastic process $Y_{t}, t \in\left[0,2 \pi\left[\right.\right.$, with $\mathbf{E}\left\{\mathrm{Y}_{\mathrm{t}}\right\}=\mathbf{E}\left\{\mathrm{Y}_{\mathrm{t}} \mathrm{Y}_{\mathrm{t}^{\prime}}\right\}=0$ and with covariance function $R(t):=\mathbf{E}\left\{Y_{0} \bar{Y}_{t}\right\}$ given by the right side of (3.4) solves

$$
\left\{\begin{array}{l}
d Y_{t}=d W_{t}-\frac{2 Y_{0}+4 Y_{t}}{2 \pi-t} d t-\frac{6 d t}{(2 \pi-t)^{2}} \int_{0}^{\mathrm{t}} \mathrm{Y}_{\tau} \mathrm{d} \tau, \quad 0 \leq \mathrm{t}<2 \pi  \tag{4.10}\\
\mathrm{Y}_{0}=\text { complex Gaussian r.v. with } \mathbf{E}\left\{\mathrm{Y}_{0}\right\}=\mathbf{E}\left\{\mathrm{Y}_{0}^{2}\right\}=0 \\
\quad \text { and } \mathbf{E}\left\{\mathrm{Y}_{0} \overline{\left.\mathrm{Y}_{0}\right\}}\right\}=\frac{\pi}{6}
\end{array}\right.
$$

driven by complex Brownian motion $W$ of intensity $d W_{t} \overline{d W}_{t}=d t$, which is independent of the initial value $\mathrm{Y}_{0}$.

Remark 4.6. It is possible to write $Y$ as an explicit linear functional of W by observing that $\left[\mathrm{U}_{\mathrm{t}}, \mathrm{Y}_{\mathrm{t}}, \mathrm{V}_{\mathrm{t}}\right]^{\top} \equiv\left[\mathrm{Y}_{0}, \mathrm{Y}_{\mathrm{t}}, \int_{0}^{\mathrm{t}} \mathrm{Y}_{\mathrm{s}} \mathrm{ds}\right]^{\top}$ is Markov and solves

$$
d\left[\begin{array}{c}
U_{t} \\
Y_{t} \\
V_{t}
\end{array}\right]=\left[\begin{array}{ccc}
0 & 0 & 0 \\
\frac{-2}{2 \pi-t} & \frac{-4}{2 \pi-t} & \frac{-6}{(2 \pi-t)^{2}} \\
0 & 1 & 0
\end{array}\right]\left[\begin{array}{c}
U_{t} \\
Y_{t} \\
V_{t}
\end{array}\right] d t+\left[\begin{array}{c}
0 \\
d W_{t} \\
0
\end{array}\right]
$$

with initial data $\left[U_{0}, Y_{0}, V_{0}\right]^{\top}=\left[Y_{0}, Y_{0}, 0\right]^{\top}$. This leads to a somewhat cumbersome (W-nonanticipative) representation of Y. It is not hard to check that the following (W-anticipative) representation also holds

$$
\mathrm{Y}_{\mathrm{t}}=\mathrm{W}_{\mathrm{t}}-\frac{\mathrm{t}+\pi}{2 \pi} \mathrm{~W}_{2 \pi}+\frac{1}{2 \pi} \int_{0}^{2 \pi} \tau \mathrm{~d} \mathrm{~W}_{\tau}, \quad \mathrm{t} \in[0,2 \pi[
$$

5. Periodic Ornstein-Uhlenbeck processes viewed as velocity processes driven by circular white noise. By the circular whitenoise process we simply mean the process $Y_{t}^{\prime}, t \in \mathbf{R}$, the formal derivative of the process Y constructed in the second part of Section 3 in the case $\alpha=0$. Now consider the periodic velocity process $Z_{t}, t \in \mathbf{R}$, which solves

$$
\begin{equation*}
d Z_{t}+m Z_{t} d t=d Y_{t}, \quad t \in \mathbf{R}, m>0, \tag{5.1}
\end{equation*}
$$

with boundary condition $Z_{0}=Z_{2 \pi}$. This means that

$$
Z_{t}=e^{-m t}\left[\int_{0}^{t} e^{m s} d Y_{s}+Z_{0}\right], \quad t \in[0,2 \pi[,
$$

with

$$
\mathrm{Z}_{0}=\frac{\int_{0}^{2 \pi} \mathrm{e}^{(\mathrm{s}-2 \pi) \mathrm{m}} \mathrm{~d} \mathrm{Y}_{\mathrm{s}}}{1-\mathrm{e}^{-2 \pi \mathrm{~m}}}
$$

Obviously, $Z_{t}, t \in \mathbf{R}$, is stationary with (discrete) spectral measure

$$
\sigma(\mathrm{d} \lambda)=\sum_{\mathrm{k} \in \mathbf{Z} \backslash\{0\}} \frac{1}{2 \pi} \frac{1}{\lambda^{2}+\mathrm{m}^{2}} \delta_{\mathrm{k}}(\lambda) \mathrm{d} \lambda, \quad \lambda \in \mathbf{R},
$$

obtained by multiplying

$$
\sum_{k \in \mathbf{Z} \backslash\{0\}} \frac{1}{2 \pi} \delta_{\mathrm{k}}(\lambda) \mathrm{d} \lambda \equiv \text { the spectral measure of } Y_{\mathrm{t}}^{\prime},
$$

by the transfer function

$$
\lambda \leadsto|\mathrm{H}(\lambda)|^{2}=\mathrm{H}(\lambda) \mathrm{H}(-\lambda)=\frac{1}{\mathrm{i} \lambda+\mathrm{m}} \frac{1}{-\mathrm{i} \lambda+\mathrm{m}}=\frac{1}{\lambda^{2}+\mathrm{m}^{2}},
$$

corresponding to (5.1). We calculate the covariance $r(t)=\mathbf{E}\left\{Z_{0} \overline{Z_{t}}\right\}, t \in[0,2 \pi[$, next. On account of (4.8) we have

$$
E\left\{d Y_{s} \overline{d Y_{t}}\right\}=\delta_{0}(t-s) d t-\frac{1}{2 \pi} d s d t .
$$

Thus

$$
\begin{aligned}
r(0) & \equiv \mathbf{E}\left\{Z_{0} \overline{Z_{0}}\right\} \\
& =\frac{1}{\left(1-e^{-2 \pi m}\right)^{2}}\left[\int_{0}^{2 \pi} \mathrm{e}^{2(s-2 \pi) \mathrm{m}} \mathrm{ds}-\frac{1}{2 \pi}\left(\int_{0}^{2 \pi} \mathrm{e}^{(\mathrm{s}-2 \pi) \mathrm{m}} \mathrm{ds}\right)^{2}\right] \\
& =\frac{1+\mathrm{e}^{-2 \pi \mathrm{~m}}}{2 \mathrm{~m}\left(1-\mathrm{e}^{-2 \pi \mathrm{~m}}\right)}-\frac{1}{2 \pi \mathrm{~m}^{2}}
\end{aligned}
$$

On the other hand, (5.1) implies that

$$
\mathrm{r}^{\prime}(\mathrm{t})+\mathrm{mr}(\mathrm{t})=\frac{\mathrm{e}^{(\mathrm{t}-2 \pi \mathrm{~m})}}{1-\mathrm{e}^{-2 \pi \mathrm{~m}}}-\frac{1}{2 \pi \mathrm{~m}}
$$

from which we find that

$$
\begin{equation*}
\mathrm{r}(\mathrm{t})+\frac{1}{2 \pi \mathrm{~m}^{2}}=\frac{\mathrm{e}^{-\mathrm{tm}}+\mathrm{e}^{(\mathrm{t}-2 \pi) \mathrm{m}}}{2 \mathrm{~m}\left(1-\mathrm{e}^{-2 \pi \mathrm{~m}}\right)} \tag{5.2}
\end{equation*}
$$

Incidentally, the expression in the right side is precisely the covariance of a periodic Ornstein-Uhlenbeck process with period $2 \pi$ (see [2]). This means that if $\gamma_{0}$ is some complex Gaussian r.v. with $\mathbf{E}\left\{\gamma_{0}\right\}=\mathbf{E}\left\{\gamma_{0}^{2}\right\}=0, \mathbf{E}\left\{\gamma_{0} \overline{\gamma_{0}}\right\}=1$, which is independent of $Y_{t}, t \in \mathbf{R}$, then the process

$$
\varpi_{\mathrm{t}}:=\mathrm{Z}_{\mathrm{t}}+\frac{1}{\mathrm{~m} \sqrt{2 \pi}} \gamma_{0}, \quad \mathrm{t} \in \mathbf{R}
$$

is the one associated with a harmonic oscillator with frequency m at a nonzero temperature $\mathrm{T}=\mathrm{k} / 2 \pi$; k is the Boltzmann constant. This process obviously solves

$$
\mathrm{d} \varpi_{\mathrm{t}}+\mathrm{m} \varpi_{\mathrm{t}} \mathrm{dt}=\mathrm{d} Y_{\mathrm{t}}+\frac{\gamma_{0}}{\sqrt{2 \pi}} \mathrm{dt}, \quad \mathrm{t} \in \mathbf{R}
$$

with periodic condition $\varpi_{\mathrm{t}}=\varpi_{\mathrm{t}+2 \pi}$.
Remark 5.1. The comment in Remark 4.2 implies that $\varpi_{\mathrm{t}}, \mathrm{t} \in[0,2 \pi[$, having covariance given by (5.2), does have the two-point Markov property (which is well known). Notice also that

$$
\frac{1}{2 \pi \mathrm{~m}^{2}}=\frac{1}{2 \pi} \int_{0}^{2 \pi} \frac{\mathrm{e}^{-\mathrm{tm}}+\mathrm{e}^{(\mathrm{t}-2 \pi) \mathrm{m}}}{2 \mathrm{~m}\left(1-\mathrm{e}^{-2 \pi \mathrm{~m}}\right)} \mathrm{dt}
$$

which shows that $\int_{0}^{2 \pi} r(t) d t=0$. It is trivial to check that the covariance function of $Y_{t}, t \in[0,2 \pi[$ integrates to 0 , too.

Finally, we consider (5.1) with $d Y_{t}$ replaced by $d X_{t}$, the circular white noise process with nontrivial holonomy described in the first part of Section 3 -the process from (4.5). Now there is precisely one $\alpha$-periodic solution (i.e., solution for which $Z_{2 \pi}=\mathrm{e}^{\mathrm{i} \alpha} \mathrm{Z}_{0}$ ):

$$
Z_{t}=e^{-m t}\left[\int_{0}^{t} e^{m s} d X_{s}+Z_{0}\right], \quad t \in[0,2 \pi[
$$

with

$$
\mathrm{Z}_{0}=\frac{\int_{0}^{2 \pi} \mathrm{e}^{(\mathrm{s}-2 \pi) \mathrm{m}} \mathrm{dX}}{\mathrm{e}_{\mathrm{s}}}-
$$

On the other hand (4.1) implies that $\mathbf{E}\left\{\mathrm{dX}_{\mathrm{t}} \overline{\mathrm{dX}}\right\}=\mathrm{dt}$ and we find that

$$
\begin{aligned}
r(0) & \equiv \mathbf{E}\left\{\mathrm{Z}_{0} \overline{\mathrm{Z}_{0}}\right\}=\frac{\int_{0}^{2 \pi} \mathrm{e}^{2(\mathrm{~s}-2 \pi) \mathrm{m}} \mathrm{ds}}{\left(\mathrm{e}^{\mathrm{i} \alpha}-\mathrm{e}^{-2 \pi \mathrm{~m}}\right)\left(\mathrm{e}^{-\mathrm{i} \alpha}-\mathrm{e}^{-2 \pi \mathrm{~m}}\right)} \\
& =\frac{1}{2 \mathrm{~m}} \frac{\sinh (2 \pi \mathrm{~m})}{\cosh (2 \pi \mathrm{~m})-\cos (\alpha)}
\end{aligned}
$$

At the same time,

$$
\mathbf{E}\left\{Z_{0} \overline{d X_{t}}\right\}=\frac{e^{(t-2 \pi) m}}{e^{i \alpha}-e^{-2 \pi m}} d t
$$

which implies

$$
\mathrm{r}^{\prime}(\mathrm{t})+\mathrm{mr}(\mathrm{t})=\frac{\mathrm{e}^{(\mathrm{t}-2 \pi) \mathrm{m}}}{\mathrm{e}^{\mathrm{i} \alpha}-\mathrm{e}^{-2 \pi \mathrm{~m}}}
$$

and we thus find that

$$
\mathrm{r}(\mathrm{t})=\frac{1}{2 \mathrm{~m}} \frac{\sinh (\mathrm{mt})\left(\mathrm{e}^{-\mathrm{i} \alpha}-\mathrm{e}^{-2 \pi \mathrm{~m}}\right)+\mathrm{e}^{-\mathrm{mt}} \sinh (2 \pi \mathrm{~m})}{\cosh (2 \pi \mathrm{~m})-\cos (\alpha)} .
$$

$J$ ust as before, since $r^{\prime \prime}=$ const $\times r$, we conclude that this covariance function corresponds to a reciprocal process-the $\alpha$-periodic (i.e., with nontrivial holonomy) version of a harmonic oscillator at nonzero temperature.

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